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THE INFLUENCE OF X-RAY TREATMENT ON FLOWER BULBS [1]) (HYACINTHS AND TULIPS)

by W. E. DE MOL.

Summary. A description is given of tests [2]) on genetic phenomena and improvement of flower bulbs, and particularly on the temporary and permanent changes brought about by X-ray treatment.

Cell division

For the assistance of the reader we shall begin with an explanation of several technical botanical terms and ideas relating to the multiplication of plant cells by division. We must distinguish between ordinary cells, those from which leaves, perianth leaves, stalks and roots are built up (the so-called vegetative or somatic cells), and cells which serve for reproduction, namely the male (pollen grains) and the female (embryo cells or germ cells) collectively called generative cells, sexual cells or gametes.

The somatic cells normally contain a constant number of longitudinal bodies, chromosomes, in their nuclei. The division (mitosis) takes place at a certain moment when all the chromosomes of a somatic nucleus split in two lengthwise (separation) and then the two products of the division (*fig. 1a*) move away from each other toward two opposite points (poles) in the cell (disjunction) to form two new cell nuclei in this way, each of which has the same number of chromosomes as the original cell before the division. This manner of division of the somatic cells is called typic mitosis or equivalent division because it gives exactly equivalent division products in the longitudinal halves of the chromosomes.

The division of sexual cells takes place in a different way. A cell which is destined to form generative cells (mother cell) undergoes a spontaneous process before the division in which the chromosomes group themselves in pairs. From each of these twin chromosomes one moves toward the one pole and the other toward the other pole. (*fig. 1b*). The result is two new cell nuclei, each of which has half the number of chromosomes of the original cell. In this case one speaks of reduction division or of allotypic mitosis. The reduction division proper (first meiotic or heterotypic division) is followed as a rule by an ordinary division (second meiotic or homotypic division),

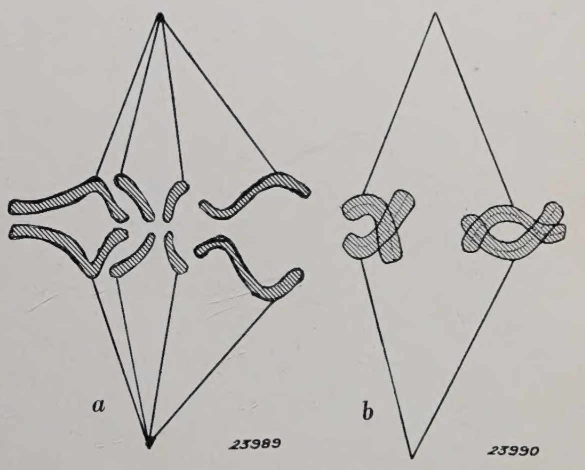

Fig. 1*a*. Ordinary division (typic division) of a nucleus with 4 chromosomes. All 4 chromosomes are split lengthwise.
Fig. 1*b*. Heterotypic division. The chromosomes are divided into 2 equivalent groups of two each. After twin chromosomes have been formed, they separate again, 2 chromosomes toward one pole and 2 toward the other.

1) *Note by the editor:* In this article a new application of X-ray treatment is described. Dr. W. E. de Mol, who has carried out investigations on this subject, has consented to explain some of his results to our readers.

so that from one mother cell four cells are formed, each of which has half the normal number of chromosomes. In ordinary cases all of these cells develop into pollen grains in the anther. In the ovulum however from each four only one germ cell is formed. Further it must be kept in mind that in fertilization a male and a female nucleus are fused together to give a germ nucleus, which then again contains the original number of chromosomes. With the surrounding plasma the germ cell is thus built up, and is the first somatic cell of the newly formed individual.

Structure and development of the bulb

In order to understand phenomena occurring in a bulb it must be remembered that a bulb may be compared with a shoot. The leaves of the shoot correspond with the scales of the bulb. At the end of the shoot or in the axil of a "leaf" the future plant develops with its leaves and flowers. This development, during which the plant is still completely hidden inside the bulb, takes place in the plants under consideration (hyacinth and tulip), between June and November. Those parts which will be situated lowest on the plant are the first to be developed.

Just as in the axils of the leaves of deciduous trees buds are formed which will grow in the following year to new shoots, so bud formation also takes place in the bulb. These buds (bulbils) which are formed in the axils of the thick scales (*fig. 2*) are the bulbs for the following year or later. One bulb can in this way produce many buds, and so many new bulbs (asexual reproduction, vegetative propagation). By vegetative reproduction therefore, an extensive "family" can be formed from one bulb. When all the members of the "family" are alike, which is normally the case, one speaks of a clone. In addition to vegetative reproduction with bulbs, there is also reproduction by means of seeds, after previous pollination, self-pollination or crossing (sexual reproduction).

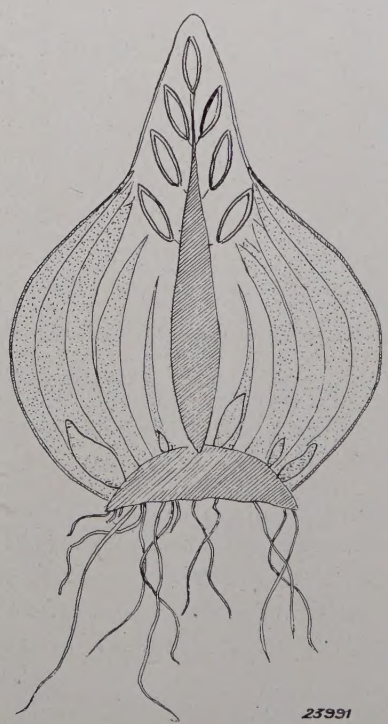

Fig. 2. Longitudinal cross section through a hyacinth bulb in the winter. Young flower stalk and bulbils clearly visible.

Preliminary tests

From 1908 up to the present time tests have been carried out by the writer on genetic phenomena and improvement, especially in the case of flower bulbs.

In the early years (1908 - 1914) attention was devoted exclusively to changes in the cells of leaves and perianth leaves. It appeared that not only modifications in the shape of the leaves and perianth leaves occurred, but also changes appeared in the colour of these organs. Many of these anomalies were found to be hereditary, so that we were concerned with somatic mutations. Up to 1914 the genetic anomalies were multiplied exclusively vegetatively. After that year the fact was also established by crossing that the deviations observed (of shape and colour) were hereditary.

From 1918 onwards the changes in the cells were studied with the help of the microscope (cytologic research). The first fact established was that in cultivation [2]) the number of chromosomes changes (increases) in many varieties. The wild hyacinth (Hyacinthus orientalis), the tulip (Tulipa Gesneriana) and the narcissus (Narcissus pseudonarcissus and poeticus) have respectively 16, 24 and 14 chromosomes in the somatic nuclei. In the cultivated form the following numbers usually appear: 24 chromosomes ($8+2\times 8$) in the hyacinth, 36 chromosomes ($12+2\times 12$) in the tulip and 21 chromosomes ($7+2\times 7$) or 28 chromosomes ($2\times 7 + 2\times 7$) in the narcissus. Other numbers of chromosomes are also found.

The probable explanation is that sometimes no reduction division takes place in the formation of the gametes, so that generative nuclei occur with the ordinary somatic number of chromosomes; thus with the hyacinth: $2\times 8 = 16$, with the tulip: $2\times 12 = 24$, and with the narcissus: $2\times 7 = 14$.

It was found that such a "doubling of the

[2]) By culture is meant the whole treatment applied by the grower for the multiplication of bulbs.

generative nuclei" can be produced by exposure to a certain temperature (cold, heat).

From pollination tests with the pollen of the hyacinth, among others, which contained, in addition to ordinary pollen grains, such "double" pollen grains, and from the cytologic examination of the germ plants produced with this pollen it was found that in addition to germ plants with the normal number of 16 chromosomes, plants with 24 chromosomes also occurred. It was concluded from this test that pollination with double pollen grains often leads to an increase in the number of chromosomes.

When plants occur with the same number of cells — and that is usually the case — those having 24 chromosomes are bigger than those with 16 chromosomes; often indeed very much bigger. In this respect the deliberate generation of double pollen grains, which the writer has been able to bring about with various kinds of plants, may be considered of great importance in practice, since in this way specimens may occur which are more robust than the original wild or cultivated variety.

Influence of X-ray treatment on the reduction division

With the ordinary garden tulip, Tulipa Gesneriana, as the object of experiment, it was found that "double" and even "quadruple" pollen grains may easily occur as a result of treatment with X-rays. While normally one parent pollen cell produces four pollen grains with 12 chromosomes, under abnormal conditions it produces two pollen grains with nuclei containing 24 chromosomes. In addition numerous other aberrations appear. Thus one or more chromosomes may be "retarded" in their passage toward the poles. They may also be eliminated. Such phenomena promote the occurrence of nuclei with differing, abnormal numbers of chromosomes. The cells with such nuclei, however, often die.

The technique of the irradiation (*fig. 3*) is as follows, according to the observations made with a variety of tulip which was studied exhaustively. The irradiation is carried out between June and November, with doses varying from 100 to 1200 *r* [3]). In June the formation of the new flower begins in the newly formed bulb. In November (or earlier) this is complete including the generative cells. The first phase of the process of growth, cell division, has then taken place. The second phase, cell enlargement, has yet to begin.

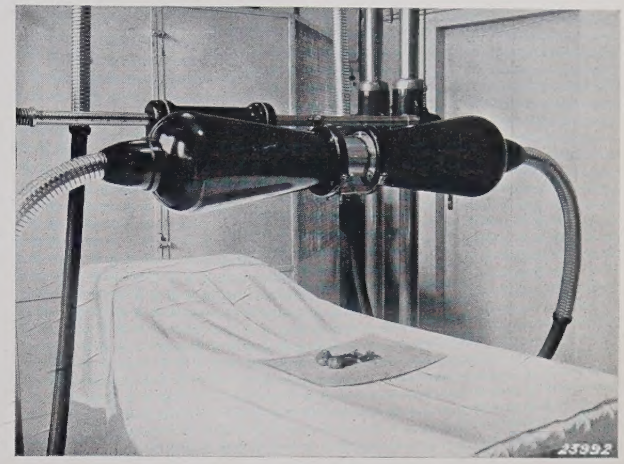

Fig. 3. Arrangement used in the X-ray treatment of flower bulbs.

Influence of X-rays treatment on somatic division

Not only allotypic mitoses but also the typical mitoses are sensitive to X-ray irradiation. It was found that with the hyacinth it was possible

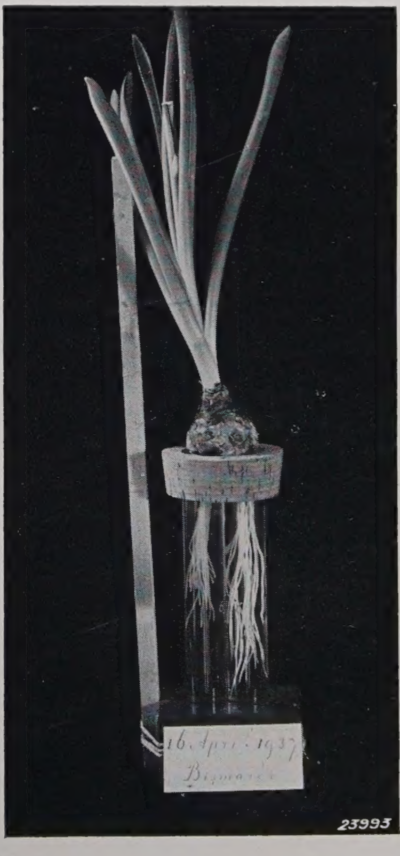

Fig. 4. Bulb of a hyacinth on a "twin glass". The roots which have developed in the right-hand glass have not been irradiated; those in the left-hand glass have been irradiated (thus the shorter roots are the irradiated ones).

[3]) An irradiation represents a dose of 1 *r* when in 1 cm^3 of air at the spot where the object is placed 0.108 erg would be absorbed as a result of the same irradiation. This corresponds to 100 ergs in 1 cm^3 of water.

deliberately to cause the formation of somatic cells with a modified nuclear structure, which are able to divide further. For this purpose the bulb was placed on a so-called twin glass filled with water (*fig. 4*). It was possible in this way to divide the roots into two groups. One group was irradiated (400 r), the other was not irradiated. The irradiated and the unirradiated root tips are studied and compared with each other.

In the former case various anomalies had appeared. Two or more adjacent cells were often found with nuclei twice as large as normal. This proves in the first place that there were somatic cells in which separation of the chromosomes but not disjunction had taken place, and in the second place that such somatic cells with double the ordinary number of chromosomes afterwards underwent normal mitosis. In the irradiated root therefore there now exist groups of double cells next to normal cells (mixo-ploidy).

Excitation of somatic mutations in shape and colour

Considering the fact that one can exert influence on the cells during their division, it is obvious that the earlier one irradiates during the process of cell division of the flower formation, the more one can influence the organs occurring lower on the stalk which are formed; thus first the leaves, then the perianth, etc.

In the first place unusual forms of perianth sometimes occur with the tulip, (*fig. 5*) as a result of irradiation. The perianth leaf becomes pointed (right), whereas it was originally rounded at the top (left). The perianth leaf (*fig. 6*) exhibits indentations on its edge and is curled to some extent (right), while these distinguishing marks were originally lacking (left). These changes have occurred because a group of cells produced from a given cell have died off, or have been enlarged so that they no longer fit into the normal cell tissue and often cause stresses in it.

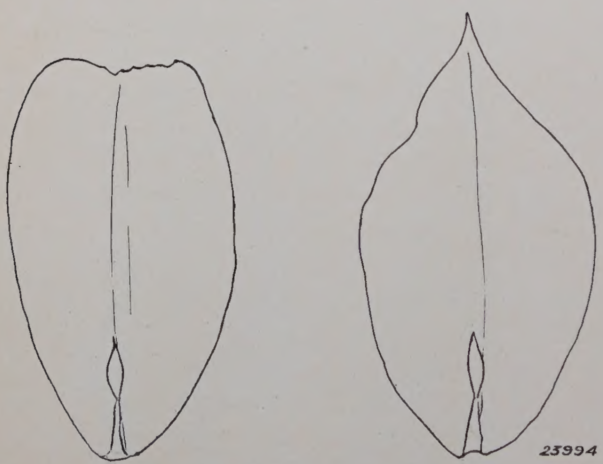

Fig. 5. Perianth leaves of the tulip Generaal de Wet. The left-hand leaf shows the original shape, the right-hand one, the shape as altered by irradiation.

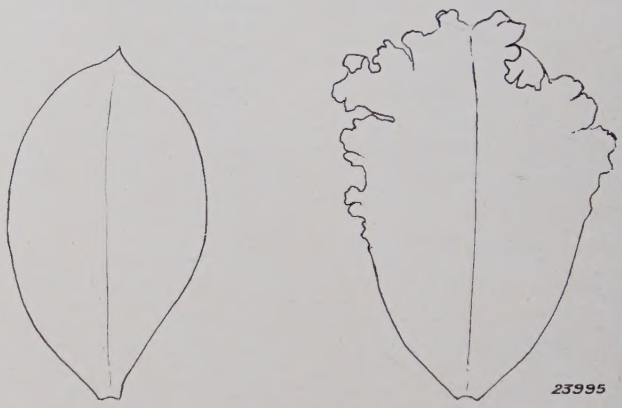

Fig. 6. Perianth leaves of the tulip Van der Neer. Left, the leaf with the original shape; right, the leaf as altered by irradiation.

At the same time, however, mutative phenomena, relating to coloration, occur in the soma. On the first occasion of their appearance they manifest themselves in a sector of a whole plant, in a sector of a whole flower or in a sector of one or more perianth leaves. The last mentioned occurs by far most frequently.

In the hyacinth a somatic mutation has remained constant (variety Grand Maître), in which three perianth leaves of the innermost corolla have changed into organs which more or less resemble anthers (*fig. 7*). On the edge, to the left and right, numerous pollen grains are developed (*fig. 8*).

Fig. 7. Two flowers of the hyacinth Grand Maître. The shape of the 3 perianth leaves of the inner whorl is altered by irradiation, and which have become stamen-like.

Moreover with the hyacinth since 1930 the rose-violet colour has been maintained after vegetative multiplication; this colour has taken the place of the blue-violet colour (also variety Grand Maître). While the blue-violet anthocyanin, as in the wild hyacinth, occurs in the lowest layer of cells of the skin of the perianth (subepidermis), the rose-violet anthocyanin is encountered in the top layer of the skin (epidermis).

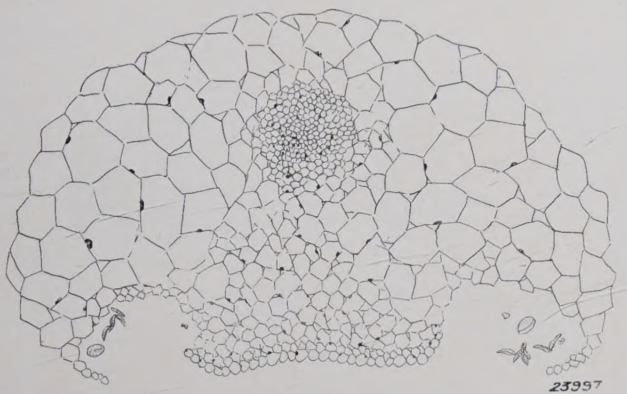

Fig. 8. Cross section of a mutated perianth leaf of the hyacinth of fig. 7. The drawing was made from a microtome preparation. It may be seen clearly that two loculi with pollen grains have developed as in the anther (lower left and right).

The following is a short description of the results of irradiation obtained in the spring of 1937, with respect to the excitation of somatic mutation in tulips.

Numerous somatic mutations in colour have occurred particularly in tulips. This year a large number of plants irradiated in 1928 bloomed in the experimental garden in Lisse. In addition the results could be seen of experiments carried out, in the majority of cases in 1933 and 1934.

We give below a schematic survey of the varieties of which one or more specimens bloomed this spring, and where it could be definitely ascertained that the flower has taken on a different colour or a different form. In this survey no statements are included about colour or shape change which up to now has been manifested only in larger of smaller sectors, or about cases in which further cultivation is necessary to determine definitely whether we are concerned with mutation (permanent change) or with modification (temporary deviation).

When it is kept in mind that for the last-mentioned tests (1933, 1934) not more than 8 or 16 bulbs of each variety were used, it is the more striking how large the chance of mutation is; at least when the experiment is well planned and well carried out.

The genetic change is not limited to a change of colour or shape of the flower. Changes of a permanent nature are also encountered in the leaves (variegated leaves). With the varieties Generaal de Wet, Herodiade, Oranje Nassau, Peach Blossom, Prince of Austria and Triumphator this phenomenon has appeared in various ways: white, yellow or yellow-green on the edges or in the middle.

In the case of the flower bulbs therefore it is the somatic mutations which are stimulated in such a surprising way by irradiation. First there is a small sector of a different colour on one perianth leaf (*fig. 9*, left), then half of a perianth leaf is of a different colour (fig. 9 middle). It is obvious that in these cases usually the corresponding portion of the stalk and of the whole plant including buds is concerned in the process of mutation, so that the following year half the flower exhibits the new colour (fig. 9 right), and finally the whole plant (*fig. 10*). This is often the progress of the development of such a vegetative mutation. This is true of course when a young bud (bulbil) has developed entirely or partially in the mutated sector.

As soon as one plant has undergone complete mutation, it may be considered as the beginning

Irradiated in 1928.

Variety	Original Colour	New Colour	Number of specimens
1. Eleonora	purple black anthers	pale violet yellow anthers	several rows (1 row = 8 specimens)
2. Roi d'Islande	pink	light pink	several plants
,,		more violet	,, ,,
,,		darker pink	,, ,,
3. William Pitt	cochineal-red	beautiful pink	14 plants, 8 flowers
,,		dark red, butterfly heart [4])	7 ,, 4 ,,
,,		dark red, blue heart [4])	50 ,, 30 ,,

[4]) By "butterfly heart" is meant that the blue spot on the inner side of the perianth leaves is bordered by a yellowish white stripe, as on the wings of a certain kind of butterfly (Vanessa antropa). If this border is lacking one sees only a "blue heart".

Irradiated in 1933 or 1934.

Variety	Original colour	New colour, one or more specimens
4. Berlin, the large-flowered mutation of Pride of Haarlem	cochineal-scarlet	a. lighter red b. still lighter red
5. Clara Butt	salmon pink	a. darker pink b. red
6. Dido	orange pink	a. lighter orange b. darker orange
7. Flamingo, Darwin	pink	a. lighter b. darker and pointed c. beautifully striped
8. Generaal de Wet	orange	a. yellow b. redder and pointed
9. General French	cherry red	light pink, pure white heart instead of yellow
10. Herodiade	pink	lighter pink
11. Ibis	underneath small portion white, further red	a. narrower red border b. white
12. Valentin	purple violet	beautifully regularly striped, no "broken colour" [5])
13. Van der Neer	purple	perianth leaves bigger and indented.

of a culture of genetically changed specimens. After several years this clone, forming a new "X-ray variety" may have increased considerably in numbers.

We have mentioned above no less than 12 different varieties of which at present one or more specimens exist with changed colour of flower, and 6 varieties with changed colour of leaf. It is clear that such somatic mutations represent a material value, sometimes a very large one, when the new colour is more beautiful than the one already existing. Quite apart from this is the scientific value of these experiments.

The hyacinths under consideration were treated with a dose of 300 *r*. The doses used for the mutation of tulips varied widely, namely from 70 to 1600 *r*.

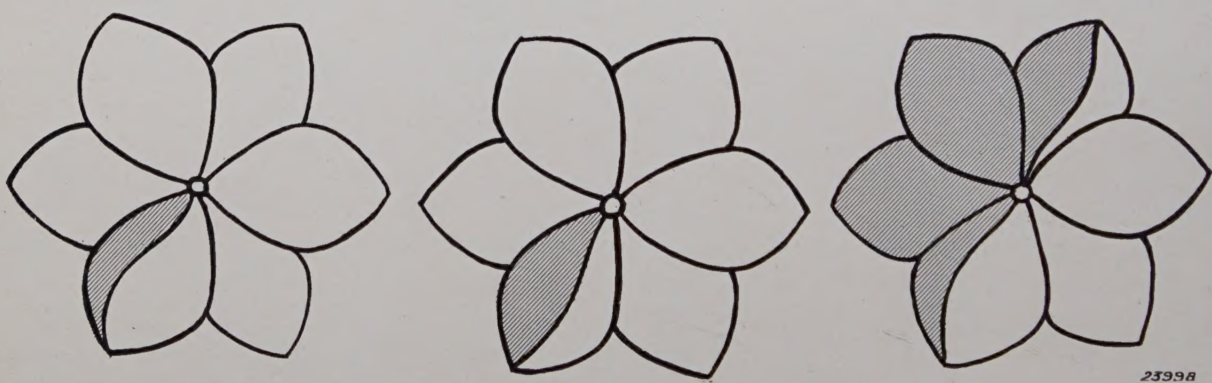

Fig. 9. Three tulip flowers drawn from below. Left: on one perianth leaf a small sector with changed colour has appeared. Middle: one leaf is half of a different colour. Right: half of the whole flower has changed its colour.

The parrot phenomenon in tulips

Extensive crossing experiments have up to now shown that the so-called parrot tulips (large perianth leaves, indented and irregular in shape and colour), cannot be multiplied by reproductive methods (by seed). In general this deviation can however be multiplied vegetatively. Considering that the genetic cells have their origin in a layer of cells situated under the outer skin (epidermis) and not in the epidermis itself, it seemed to the writer that the parrot character consists in a

[5]) When one speaks of "broken colour", one means that the original colour is broken by stripes and spots in which no anthocyanin has developed. It is a mosaic effect, perhaps caused by a virus. In the variety Valentin the new colour pattern is not due to this cause, but is a consequence of the irradiation.

genetic change exclusively in the cells of the epidermis and is therefore based upon a permanent change in the outermost cell layers of the plant. Several times "p a r r o t sectors" have been produced by X-ray treatment of a variety of tulip, but on the outside only of several perianth leaves and not on the inside. In this case this character is very

Fig. 10. From one bulb two flowerstalks have sprung. One of the flowers has entirely changed its colour, the other has retained completely the original colour. Such cases are extremely rare. In general two bulbs are formed with different coloured flowers.

probably excitated only in the epidermal cells limiting the o u t s i d e of the perianth leaf, and not in those cells which form the i n n e r s i d e of that leaf. Consequently the parrot character was not completely developed. The perianth leaf had for instance retained its normal colour on the inner side, but had taken on the parrot coloration on the outside, etc. It is obvious that this character could not be retained in this case upon further vegetative multiplication, since at least one bud must be formed from this tissue if the genetic changes are to be permanent.

This fact furnishes a new basis for the abovementioned opinion that the genetic anomalies appear during the first phase of growth, the period of cell division, and are afterwards passed on from cell to cell.

The division hypothesis

In the opinion of the writer all genetic anomalies observed, that is cells observed with an abnormal number of chromosomes, or of abnormal colour or size, can be ascribed to the action of external influences acting during the extremely subtle process of nuclear division. These influences may be the more ordinary temperature influences as well as very unusual ones, such as treatment with X-rays. These often result in the process of division, i.e. the motion of the chromosomes toward the poles being entirely prevented, retarded or accelerated. The obvious conclusion is that in many cases this intervention acts only on certain g e n e s, that is to say on the assumed material vehicles of the hereditary character (hereditary factors), either incorporated in the chromosomes, and therefore within the cell nucleus, or lying outside it. This hypothesis, proposed by the writer in 1928 and called in short "division hypothesis" will be able to guide investigators in the interpretation of the effects obtained, or in the search for new phenomena in this subject.

Literature

(Preliminary communications and short reports of lectures relating to irradiation of flower bulbs in the X-ray laboratory of the University of Amsterdam).

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No. 13 (1937). De beteekenis van eenige röntgenbestralingsproeven met planten voor de geneeskundige röntgenologie. Handel. 26e Nederl. Nat. Geneesk. Congr., 265 - 267.

Further see Herba Topiaria (Study of ornamental plants) organ of the Netherlands Society for the promotion of scientific improvement of ornamental plants (Nederlandsche Vereeniging tot bevorderiug der wetenschappelijke veredeling van siergewassen.)

These tests were enabled to be carried out in the X-ray laboratory of the University of Amsterdam, thanks to the kind collaboration of Prof. Dr. J. van Ebbenhorst Tengbergen.

The microscopical examinations took place in the laboratory for physiological chemistry of the University of Amsterdam. The writer's thanks are due to the director Prof. Dr. B. C. P. Jansen.

In the department belonging to Prof. Combes (Sorbonne) at the World Exhibition in Paris (experimental biology) extensive material of the writer was exhibited, consisting of photographs, drawings and watercolours, relating to the subject treated in this article.

NOISE IN AMPLIFIERS CONTRIBUTED BY THE VALVES.

by M. ZIEGLER.

Summary. In this article the contribution of the amplifier valves to the noise in amplifiers and receiving apparatus is discussed in some detail. We have introduced the concepts "noise factor", "noise voltage" and "noise resistance" as suitable quantities for judging amplifier valves in this respect. A special study is made of the decrease in noise due to the influence of the space charge, and of the increase in noise in screen grid valves, due to fluctuations resulting from the division of the current. Finally the principle is explained of a "noiseless" screen grid valve developed in this laboratory and the advantages of this valve over existing valves are discussed.

Introduction

In a previous article in this periodical [1]) we became acquainted with the fundamental causes of noise in amplifiers: the thermal fluctuations of the electricity in conductors and the shot effect of currents of electrons in amplifier valves. We have seen that the irregular alternating voltage, which occurs between two points in an electric network as a result of the thermal fluctuations of electricity, has been established theoretically and practically. The contribution of this phenomenon to the total noise in amplifiers and radio receiving sets can always be calculated in a simple way from the formulae derived. In this article we shall not study further the thermal fluctuations, but confine ourselves to the noise contribution of the valves.

Noise factor

As a consequence of the corpuscular nature of electricity the anode current of every amplifier valve exhibits fluctuations about the mean value, even when the potentials of the electrodes are constant. In the very simple case of the saturated diode, that is a diode in which the whole emission passes over to the anode, one is concerned with the fluctuations in time of the number of independent electron emissions; the electrons emitted and therefore collected by the anode have irregular distribution in time; this is the case of pure shot effect. In the above mentioned article we derived the following formula for the value of these fluctuations:

$$\Delta\,\overline{(I-\overline{I})^2} = 2\,e\,\overline{I}\,\Delta\,\nu, \quad \cdots\cdots (1)$$

in which the left-hand member of the equation represents the contribution to the mean square of the fluctuations of the current, which is due to the components included in the frequency interval between ν and $\nu + \Delta\,\nu$.

If we are not considering a saturated diode but an amplifier valve, the situation is not so simple, and the fluctuations of the anode current have a different value from the one given by (1). This can be expressed by introducing into the formula for the fluctuations of a current of electrons a factor usually indicated by F^2.

$$\Delta\,\overline{(I-\overline{I})^2} = F^2\,2\,e\,\overline{I}\,\Delta\,\nu. \quad \cdots (2)$$

We shall call this factor F^2 the **noise factor**.

By the noise factor of a current, therefore, we mean the ratio of the mean square of the fluctuations of the current to the mean square of the fluctuations of another current which is on the average equally large and exhibits pure shot effect.

Noise voltage and noise resistance

In the course of this article we shall be concerned. with the factors which determine the magnitude of F^2. The first problem is the determination of the degree to which the fluctuations of the anode current of a valve act as disturbances in the reception of a signal.

In order to obtain some idea of this, these irregular fluctuations of the anode current of an amplifier valve may be compared with the variations of the anode current which are caused by a small voltage introduced between the cathode and the control grid of the valve. It is clear that the smaller the alternating grid voltage which causes current fluctuations whose mean square is equal to that of the fluctuations which are naturally present, the less disturbing these latter undesired fluctuations will be. The alternating grid voltage, which is called the **equivalent noise voltage** of the amplifier valve, is therefore an important factor.

[1]) The Causes of Noise in Amplifiers, Philips Techn. Rev. **2,** 136 (1937).

The noise voltage is found by dividing the value of the fluctuations of the anode current by the slope of the valve.

$$\Delta\,\overline{V^2} = \frac{\Delta\,\overline{(I-\bar{I})^2}}{S^2} = \frac{F^2\,2\,e\,\bar{I}}{S^2}\,\Delta\,\nu \quad \cdot \cdot \; (3)$$

One may calculate further the magnitude of the resistance which exhibits thermal fluctuations of voltage between its extremities equal in size to those of (3). In the article mentioned we found that the thermal fluctuations in voltage of a resistance R at the absolute temperature T are given by

$$\Delta\,\overline{V^2} = 4\,k\,T\,R\,\Delta\,\nu, \quad \cdot \cdot \cdot \cdot \; (4)$$

in which k is the Boltzmann constant.

When (3) and (4) are combined one obtains

$$R = F^2\,\frac{2\,e\,\bar{I}}{4\,S^2\,k\,T} \quad \cdot \cdot \cdot \cdot \cdot \; (5)$$

With $e = 1.60\times10^{-19}$ coulomb, $k = 1.37\times10^{-23}$ erg/degree and $T = 290°$ K, R is found to equal 20 000 $F^2\bar{I}/S^2$ ohms, when $\bar{I}$ is in mA and S in mA/Volt.

This value of R, which is called the equivalent noise resistance of the amplifier valve is a useful measure of the quality of a valve with respect to noise. According to the definition of R the mean square of the fluctuations of the anode current is exactly equal to the thermal fluctuations which occur when the impedance between grid and cathode at the frequency under consideration is equal to R. When R is known, one immediately knows for any arbitrary circuit which source of noise furnishes the largest contribution; the amplifier valve or the electric circuits between grid and cathode.

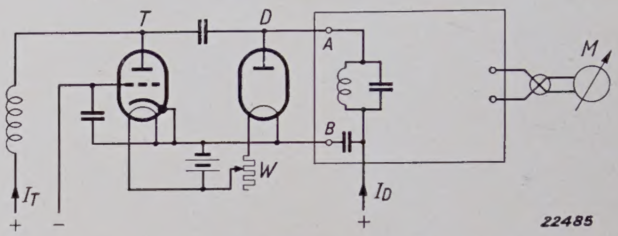
The determination of the noise resistance of an amplifier valve involves measurement of the noise factor F^2 with which we shall now deal.

Measurement of the noise factor

In order to determine directly the noise factor of a current with fluctuations of unknown intensity, one might use the measuring arrangement which was described in the article already cited on the causes of noise, and calculate the factor F^2 from the result of the measurement with the aid of equation (2).

A simpler method is the following. The well defined fluctuations of the current of a saturated diode are used as a standard to measure the unknown fluctuations. The arrangement used for this purpose is given diagrammatically in *fig. 1.*

T is the amplifier valve for which the fluctuations of the current to one or more electrodes must be measured. These fluctuations are amplified with a selective and linear amplifier which passes a certain band of the frequency spectrum so narrow, that it may be assumed that when the



Arrangement for the determination of the noise factor of the anode current of an amplifier valve T. By varying W, the anode current from the saturated diode D is adjusted so that the noise contribution of D is equal to that of T. The noise factor F^2 is then equal to I_D/I_T.

factor F^2 is dependent on frequency it may be regarded as practically constant in the frequency range transmitted. The amplified fluctuations are measured in an instrument M which indicates the mean square of the value of the output current.

In parallel with T, *via* a condenser of large capacity, is connected a diode D having a tungsten filament, whose temperature can be adjusted to any value by means of the filament control resistance W. The anode voltage of the diode is so high that all the electrons emitted are collected, and the fluctuations of this current therefore correspond with the pure shot effect. Furthermore, the internal resistance of this diode is always very high, so that the impedance between points A and B and thus the amplification of the fluctuations of T, is independent of the current adjustment of the diode.

If the amplification is adjusted for a certain value of the current I_T through T and $I_D = 0$ so that the meter M has a definite deflection, one may then, with the aid of W, regulate the current I_D so that the indications of M are doubled. Then the fluctuations of I_D are equally as large as those of I_T, since they are entirely independent of each other and therefore their mean squares may be added.

According to the definition given in the previous section F^2 is then exactly I_D/I_T.

When the fluctuations of the currents of various kinds of amplifier valves are measured in this way, very different values are found.

In general values will be found which are smaller than 1, but values larger than 1 will be found. The former values indicate that a certain regularity has appeared in the current of electrons which passes over to the electrode under consideration,

and we shall see in the following that the space charge in an amplifier tube decreases the disorder in the transport of electrons to the anode, and therefore also decreases the factor F^2. The second case, $F^2>1$, is an indication that there must be another source of fluctuations besides the complete disorder with which the electrons are emitted. With complete disorder F^2 has the value 1. Only a quite different source of fluctuations such as a kind of abnormal modulation of the emission. due to temporary changes in the state of the cathode, the presence of other charged particles beside the electrons, or secondary emission, can provide an explanation of this.

The influence of secondary emission may be conceived as follows. Primary electrons which strike an electrode with sufficient speed may at the same time free more than one, for instance n secondary electrons: one speaks then of multiple secondary electrons which may be conceived as a charged particle with a charge equal to n times the charge on an electron. A current consisting of particles with the charge $n\,e$ must, according to the formula for the shot effect, exhibit fluctuations which are n times larger than the fluctuations of single electrons [2]).

Fluctuations in the state of the cathode itself may be caused by the presence of foreign particles which diffuse to the surface and cause local and temporary changes in the emission [3]). This phenomenon has nothing in common with the shot effect. The most striking difference is the dependence on frequency: these fluctuations are relatively slow and decrease sharply in intensity at high frequencies, in contrast to the shot effect fluctuations which, as we have seen in the article mentioned, are distributed uniformly over the whole frequency spectrum. At low frequencies the fluctuations due to the state of the cathode may exceed the shot effect in intensity by a large factor.

The presence of ions from the ionization of gas residues, or the emission of ions from the cathode may also lead to abnormal fluctuations [4]).

These phenomena however are not inherent in the principle of the amplifier valve, and need not necessarily appear in practice. We shall therefore further confine ourselves to the simple case in which the cathode (and only the cathode) emits electrons (and nothing else but electrons) and shows no changes with time, and the electron current is influenced by nothing else but the presence of the electrons themselves together with the constant voltage on the electrodes.

Influence of the space charge on the noise factor

If in our arrangement for the measurement of the noise factor, T is a saturated diode which satisfies all the conditions mentioned above (the standard diode D of course has been rigorously tested in this respect), we always find $I_D = I_T$, i.e. $F^2 = 1$. In a triode however in which the anode current is only a fraction of the emission, as a consequence of the presence of a cloud of electrons (space charge) around the cathode, a potential minimum occurs between anode and cathode, so that only the speediest electrons reach the anode. It is found that a change in the emission has an effect on the depth of the minimum, such that the number of electrons transported per second remains practically the same.

Fluctuations in the emission will therefore in this case be reproduced to a much smaller degree in the electron current to the anode, than when this is the saturation current. There appears therefore a decrease in the fluctuations due to the space charge; $F^2<1$. This decrease is a function of the saturation current I_s, the anode current I_a, and the geometry of the valve. For a given valve at a certain value of I_s one may determine the noise factor as a function of I_a/I_s. It is found that with decreasing value of I_a/I_s the noise factor is at first reduced. The decrease of F^2 with I_a/I_s may be very appreciable, but it is not unlimited. It is clear that, if by decrease of the anode voltage the anode current is made continually smaller, a state is finally reached at which a potential minimum no longer exists, but the potential of the anode is the lowest. In that case the space charge between cathode and anode has no influence on the current strength: all the electrons which have sufficient speed of emission to move against the negative anode voltage will reach the anode, and the current strength will fluctuate just as sharply as the number of these electrons. The emission of electrons with a speed higher than a definite value may, like the total emission, be considered as a series of independent events. The anode current thus again exhibits chance fluctuations and $F^2 = 1$. Thus F^2 as a function of I_a/I_s has a minimum. In *fig. 2* the curve for the noise factor is represented for a diode adjusted at an emission of 4.5 mA.

As an illustration of the decrease of fluctuations due to the space charge, we should like to mention measurements carried out on pentode EF 5, which was connected as a triode in this experiment by joining screen grid and plate and considering them as one anode. At the following adjustment:

$$V_{a+g} = 100\text{ V},\ \ V_{g_1} = -2.5\text{ V},\ \ I_{a+g} = 10\text{ mA}$$

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[2]) M. Ziegler, Shot Effect of Secondary Emission, Physica, **3,** 1 and 306 (1936).

[3]) W. Schottky, Small Shot Effect and Flicker Effect, Phys. Rev. **28,** 75 (1926).

[4]) S. Ballantine, Fluctuation Noise due to Collision Ionisation in Electronic Amplifier Tubes, Physics, **4,** 294 (1933).

with the measuring arrangement described above we found:

$$I_D = 0.5 \text{ mA, thus } F^2 = I_D/I_T = 0.05$$

Since with this adjustment the slope of the triode is 2.2 mA/V, we calculate for this case a noise resistance of about 2000 ohms.

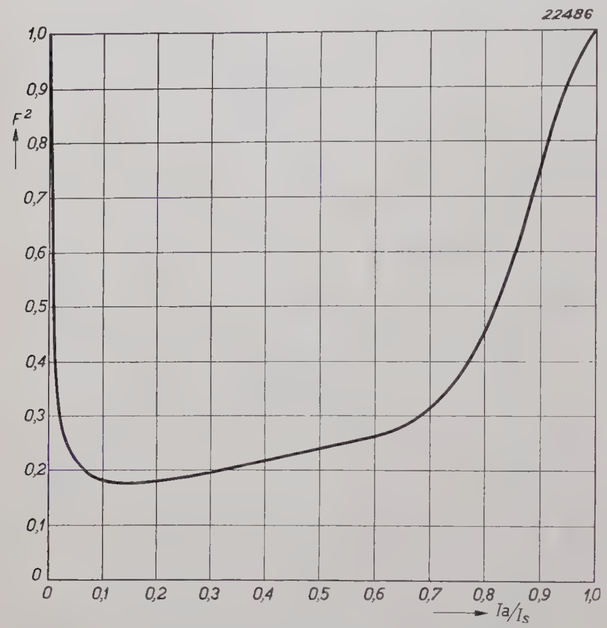

Fig. 2. Curve of the noise factor as a function of the current at constant emission from a diode with a tungsten filament.

As we have already seen this means, that in the presence of a resistance of 2000 ohms between grid and cathode, the thermal voltage fluctuations in this resistance cause anode current fluctuations which are as large as the fluctuation of the anode current itself. Since in general the resistance, i.e. the real portion of the impedance present between grid and cathode, in the frequency range under consideration, is much greater than 2000 ohms, the contribution thereof to the total fluctuations will be large compared with the contribution due to the amplifier valve mentioned.

Noise due to division of the current

It is known that because of the capacity between screen grid and anode a triode is not suitable for the amplification of voltages of high frequency [5]). Therefore a screen grid kept at a constant potential has been introduced between grid and anode, which makes the capacity very small. This however involves the fact that the total cathode current is divided into anode current and screen grid current. The quality of the valve is determined by the noise factor of the anode current: only the fluctuations of the anode current are superposed on the desired anode current variations which are caused by alternating voltage of the control grid to be amplified. If now the noise factor of the pentode EF 5 connected in a normal way is measured at the adjustment $V_{g_1} = -2.5$ V, $I_a = 7.5$ mA, $I_a + I_g = 10$ mA, which corresponds exactly with the adjustment given above for the same valve as triode, one finds $F_a^2 = 0.28$.

This means that, relatively as well as absolutely, the fluctuations of the anode current are much greater than the fluctuations of the total current. This may be ascribed to the appearance of the so-called division fluctuations: the total current fluctuates but little, each of its parts, however, fluctuates more strongly.

This phenomenon, like the shot effect, is due to the fact that the current consists of individual electrons.

The division of the total and practically non-fluctuating current into anode current and screen grid current depends at every moment on the chance position and direction of the electrons which are moving toward the positive electrodes: for a short moment more electrons will fall on the screen grid than correspond to the average, then again the anode will receive more electrons than the average number. How large are these fluctuations?

It is easy to calculate this for a special case. Assume that the electrons coming from the space charge follow each other at equal intervals of time so that no fluctuations appear in the total cathode current and the noise factor of the cathode current is therefore zero. Assume further that the screen grid is very widely spaced, so that by far the majority of the electrons reach the anode, or in other words that I_g is small compared with I_a. Now and again the direction and velocity of an electron will by chance be such that it must fall on the screen grid: the transitions of the electrons which go to the screen grid thus form a series of independent events, and the fluctuations of their number with the time are given by the formula for the shot effect. The fluctuations of the green grid are thus:

$$\Delta\,\overline{(I_g - \overline{I}_g)^2} = 2\,e\,I_g\,\Delta\,\nu.$$

The fluctuations of the anode current are of course equal and opposite, since the total current is strictly constant. The noise factor of the anode current is by definition therefore:

$$F_a^{\,2} = \frac{2\,e\,I_g\,\Delta\,\nu}{2\,e\,I_a\,\Delta\,\nu} = \frac{I_g}{I_a}.$$

[5]) See for example the article "Five-electrode Transmitting Valves (Pentodes)" Philips Techn. Rev. 2, 257 (1937).

For any arbitrary ratio between I_a and I_g for the case when the total current does not fluctuate and the division of the current obeys only the laws of chance, one finds the following:

$$F_a^{\ 2} = \frac{I_g}{I_a + I_g} \quad . \quad . \quad . \quad . \quad . \quad . \quad (6)$$

Upon applying equation (6) to noise caused by the division of the current we must still take into account two deviations from the above assumption:

1) In reality the noise factor F_k^2 for the total current is not zero, so that F_a^2 is larger than the value given by (6). The difference $F_a^2 - F_k^2$ however is smaller than I_g $(I_g + I_a)$, since as the total current is not entirely free from fluctuations, the influence of chance in the division of the electrons is less noticeable. The less F_k^2 differs from 1, i.e. the more complete disorder is approached in the total current, the less is the disorderly division able to bring about new disorder; F_a^2 cannot be greater than 1 (total disorder).

2) In reality the division of the electrons between the screen grid and the anode is not always purely by chance, but may obey a certain geometrical regularity. It may happen that the electrons from certain points of departure on the cathode always land on the anode, and those from other points always on the screen grid.

This would occur if the screen grid did not consist of many thin wires but of a number of platelike strips. The portion between each strip and the cathode must then be considered as an individual triode, wherein therefore no division fluctuations occur. This phenomenon may be expressed mathematically by writing instead of equation (6) the following:

$$F_a^{\ 2} = a \, \frac{I_g}{I_a + I_g},$$

in which the chance factor $a < 1$.

In the case of the amplifier valve EF 5 described, where $I_g/(I_a + I_g) = 0.25$, one is concerned with both phenomena, as may be seen clearly from the values given for F_k^2 and F_a^2.

Principle of the "noiseless" valve

We have seen that the mean square of the noise voltage and the noise resistance of an amplifier valve are proportional to the noise factor F^2, the anode current I_a and inversely proportional to the square of the slope S. In order to attain a small noise resistance therefore I_a/S^2 and F^2 must be as small as possible. I_a and S are quantities which are more or less fixed for a given type of valve by the dimensions of the cathode and the requirements placed on the shape of the I_a-V_g-characteristic. A steep slope with a small anode current and the retention of all the desired properties is a goal which has always been aimed at for other reasons. Good valves are at present at the limit of the attainable in those respects. The noise resistance of the existing high-frequency amplifier valve could therefore only be made smaller by reducing the noise factor.

We have seen that the fluctuations of the anode current of a high-frequency amplifier valve may be ascribed chiefly to division fluctuations; one must therefore try to keep these division fluctuations as small as possible. Formula (6) shows us that they are about proportional to the screen grid current.

In a high-frequency amplifier valve developed in the Philips laboratory this current has been made very low. This was achieved by introducing between control grid and screen grid an extra grid which is wound with a pitch equal to the pitch of the screen grid, and is mounted so that the wires of the screen grid seen from the cathode are just in the shadow of those of the extra grid. If this extra grid is connected to the cathode, the distribution of the field is such that the electrons are, as it were, forced to move through the meshes of the extra grid and screen grid; only a small portion can reach the screen grid. Entirely in agreement with theory, it is then found that the increase of the noise factor due to division fluctuations is very small. The following are examples of values found for such an amplifier valve:

for $I_a = 8$ mA $\qquad I_g = 0.2$ mA
$\quad F_k^2 = 0.06 \qquad F_a^2 = 0.08$.

The slope is about 2 mA/V, so that the noise resistance of this valve according to (5) is

$$R = 20000 \times 0.08 \times 8/4 = 3200 \text{ ohms.}$$

For the normal EF5 with $I_a = 7.5$ mA, $S = 1.65$ mA/V and $F_a^2 = 0.28$, $R = 15\,500$ ohms. We see therefore that by reduction of the screen grid current an considerable improvement is achieved. The use of this special valve will thus result in a noticeable decrease of noise in the case where the real portion of the impedance between control grid and cathode is of the order of several thousand ohms.

THE ADJUSTMENT OF SYNCHRONIZERS FOR THE "PHOTOFLUX" PHOTO FLASH BULB

by J. A. M. VAN LIEMPT and J. A. DE VRIEND.

Summary. A description is given of a simple method of adjusting synchronizers so that the shutter of the camera is opened just at the moment of maximum intensity of the flash-light.

A synchronizer is a small subsidiary apparatus used in combination with a flash-light lamp and a photographic camera to provide that the evolution of light by the flash-light lamp takes place at the moment when the shutter is open.

In America this apparatus is in general use and practically every press photographer is equipped with one; in Europe it has not yet come into general use. We do not doubt that the situation here will soon change since synchronizers have important advantages.

In the first place they permit the photographer to operate both the shutter and the lamp with a single movement, which is a considerable simplification.

Further they provide that the light emission of the flash-light takes place at the right moment, that is, when the shutter is open widest. Considering that the time of the flash of a flash-light lamp is usually of the order of 1/40 sec, this means that with a shutter speed of 1/40 sec the whole of the illumination of the flash is used by the camera and unwanted (for instance frontal) illumination is reduced to a minimum.

In cases where the nature of the moving object to be photographed demands a short exposure time, of 1/100 sec or less, the synchronizer (at least one of the better kinds) permits us to make use of a portion of the total light emitted, and thus makes possible the taking of good snapshots in cases where lack of sufficient daylight or artificial light would otherwise prevent this.

Since the "Photoflux" flash bulb is made with a practically constant peak time, i.e. the time which passes between switching on the lamp and the maximum light development, the synchronizer, which operates not only the switch of the flash-bulb but also the shutter, especially with such short exposure times, must provide that the movement of the shutter begins and ends at such a time that as much light as possible from the flash-light lamp is used.

We are confining this discussion to Compur shutters. With focal plane shutters, each case must be examined individually since many focal plane shutters on the market are unusable with flash-light bulbs because the time of traverse of the slit over the plate is so long that the flash would expose only a portion of the negative.

Synchronizers of various kinds may be obtained. As will appear in the following, an **adjustable** synchronizer is to be recommended in which the instant of shutter operation may be varied.

For the sake of clearness we will briefly describe one of the existing types. In this type the flash-light is ignited by switching on the current from a pocket battery. In parallel with this is an electromagnet, whose armature operates the shutter release. By variation of the distance between the armature and the core of the electromagnet, and by adjusting the movement of the armature to be quick or slow, it is possible to produce variation in the time between ignition of the flash and operation of the shutter.

When an adjustable synchronizer is to be employed it must be adjusted to suit the camera and the flash-bulb used. By means of a cathode ray tube it is possible to determine accurately not only the peak time of the flash-light bulb but also the time constant of the synchronizer at different adjustments [1]). This method, however, is too complicated for the ordinary photographer and therefore we are giving the following simple method of carrying out the desired adjustment, particularly for the case when the Philips "Photoflux" is used as flash-light.

A cardboard disc with a diameter of about 40 cm should be covered on one side with dull black paper. Round the edge of the blackened side of this cardboard, at equal distances, eight or more white, and the same number of dark grey strips of paper are glued, having the dimensions 3 × 15 mm (see *fig. 1* where the dark grey strips are shown

[1]) J. A. M. van Liempt and J. A. de Vriend, Physica 4, 703, 1937.

cross-hatched). This disc is fastened to a gramophone disc, a loose bicycle wheel, a motor shaft or the like. When a gramophone is used, which has the advantage of running with a constant speed

Fig. 1. The cardboard disc with eight white and eight grey strips.

(78 revolutions per minute), it is preferable to set it on its edge.

A portion of the circumference of the disc is then photographed in a partially darkened room with the aid of the "Photoflux" and the synchronizer, at as nearly as possible full size (that is, with the camera close to the disc) and with an *effective exposure time of 1/100 sec and a stop of F.11 or smaller*. A fixed black paper arrow set up in the direction of motion of the disc is photographed at the same time.

The picture of the strips will be quite different in appearance according as to whether the shutter was opened too early, too late, or at the correct moment. The case where the shutter was much too early or much too late, in which no picture at all is obtained, occurs very seldom. Usually in such a case the pocket battery will be found to have run down or the contacts in the circuits to have too much resistance. When these points are in order one should adjust the synchronizer until a picture is obtained.

If the shutter opening is slightly too early, a picture like that in *fig. 2* is obtained. The distinguishing feature of this picture is that the intensity increases in the direction of the arrow. If the shutter is too late, one obtains a picture like that in *fig. 4;* this is distinguished by the fact that the intensity decreases in the direction of the arrow. If the shutter is open at exactly the right moment a picture like that in fig. 3 is obtained, with uniform intensity. By regulating the synchronizer it can always be set at the proper position in this way. In connection with the fact that most synchronizers and shutters and also the flash-light bulbs are subject to small variations, it is advisable to check the correctness of the adjustment several times in order to reach a good average.

The method here described is based upon the following considerations.

The light transmission of a Compur shutter with an effective opening time of about $^1/_{100}$ and $^1/_{200}$ sec is not the same at every moment, because of the fact that the movement of the sectors takes place with a finite velocity, but it may be represented by the ordinate of *fig. 5*. The continuous line in the figure holds for the case where the diaphragm is wide open (F . 3.5). When the opening is small (F . 11

Fig. 2. Shutter too early. The blackening increases in the direction of the arrow.

Fig. 3. Shutter operating at the right moment. The blackening is uniform.

Fig. 4. Shutter too late. The blackening decreases in the direction of the arrow.

for instance) the maximum aperature is limited by the stop and is reached much sooner, as may be seen from the dotted line. The characteristic of the shutter is reduced therefore from approximately a triangle to a rectangle.

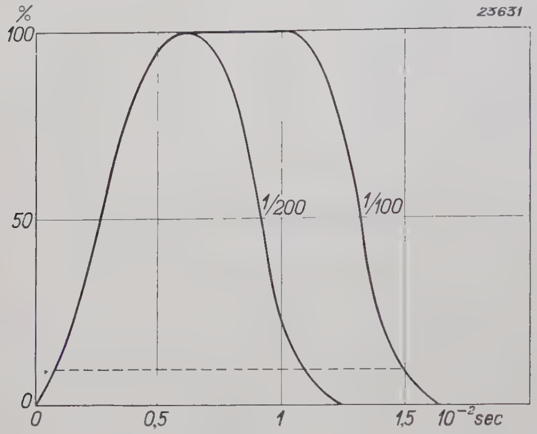

Fig. 5. The light transmission of a Compur shutter as a function of the time, for effective times of exposure of about $^1/_{100}$ and $^1/_{200}$ sec.

In *fig. 6* may be seen the flash curve of "Photoflux" Type II as it is recorded with a cathode ray tube [2]). The so-called "practical" flash time of this lamp is about $^1/_{40}$ sec and is thus shorter than the base of the curve in fig. 6, since the smaller intensities of light at the extremities give no negative blackening under the conditions under which the lamp is used, owing to the existence of a threshold value of the plate sensitivity.

The light transmission of the combined flash shutter can now easily be calculated from the

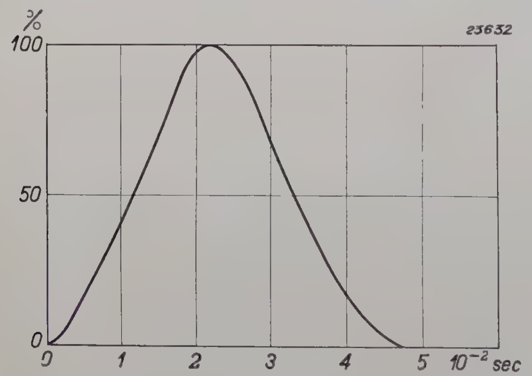

Fig. 6. The flash curve of "Photoflux" type II.

two curves (F . 11 and 1/100 sec) when the shutter opens 0.005 sec too early, or too late, or at the right moment. This is shown in *fig. 7*. From the figure it may be seen that if the shutter is too early or too late the light curve is asymmetrical with an obvious difference of intensity between the beginning and the end of the movement of the shutter, which is expressed on the negative as an easily observable difference in blackening. With greater deviations than 0.005 sec the difference is still

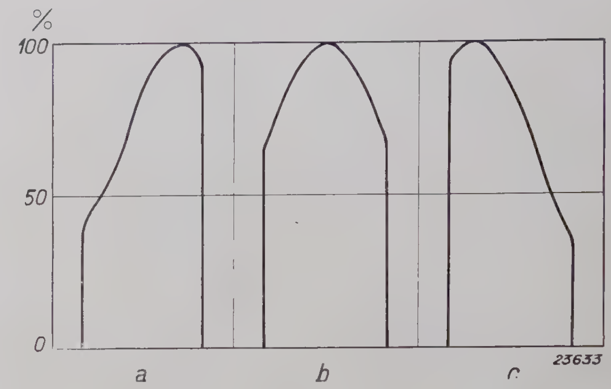

Fig. 7. The light transmission of the combination "Photoflux" (Fig. 6) with shutter $^1/_{100}$ sec and F . 11 (Fig. 5), when shutter is *a*, 0.005 sec too early, *b*, opened at the right moment and *c*, 0.005 sec too late.

more pronounced. In order to accentuate the difference in blackening the dark grey strip is glued to the disc and when the synchronizer is incorrectly adjusted, the grey strip fails in part to be shown in the picture because the brightness corresponding with its colour remains below the threshold value of the negative and thus simplifies the reading of the result. Reference is also made to fig. 2 in which the picture of the white strip after being interrupted appears again for a moment. This is caused by a common defect of the shutter consisting of the fact that after it has shut it opens slightly again and then shuts once more. This phenomenon is shown so plainly here because we were using a small opening of the diaphragm and because the light intensity in the case shown in fig. 2 was just at its maximum at the moment when the shutter reopened. The observation of this phenomenon is another proof of the fact that the shutter opened too soon.

We call attention to the fact that the "Photoflux" bulb [3]) is especially suitable for use with synchronizers because of its symmetrical somewhat flat-topped characteristic. We are concerned practically with small variations of the shutter movement caused by differences in humidity, temperature and friction, in addition there are variations in the flash-light bulb, the internal resistance of the pocket battery and also in the mechanism of the synchronizer itself.

When the synchronizer is once adjusted to an

[2]) See J. A. M. van Liempt and J. A. de Vriend, Physica **4**, 353, 1937, and Philips techn. Rev. **1**, 289, 1936.

[3]) For a detailed description see J. A. M. van Liempt and J. A. de Vriend, Philips techn. Rev. **1**, 189, 1936.

average time of retardation, these variations will never lead to failures, since the shutter always remains open in that region of the light characteristic of the lamp where the light intensity is still adequate.

With the aid of figs. 6 and 4 the light transmission can easily be construed for the combination of shutter ($^1/_{100}$ sec) and "Photoflux" for the case where the half time value of the maximum opening of the shutter coincides with the peak value of the light emission of the lamp, and for the case where the deviation from coincidence is $\pm$ 0.005 sec. This has been done for a large lens opening (in this case F . 3.5) with which these lamps are generally used, and it is shown in *fig. 8*. As may be seen, the total amount of light which is allowed to pass scarcely differs, while the flash time is also unchanged. In addition it may be seen from this figure that the above described method of adjusting a synchronizer fails when a small opening of the diaphragm is not used. Only at a stop of F . 11 or smaller are the characteristic differences between too early and too late opening of the shutter manifested.

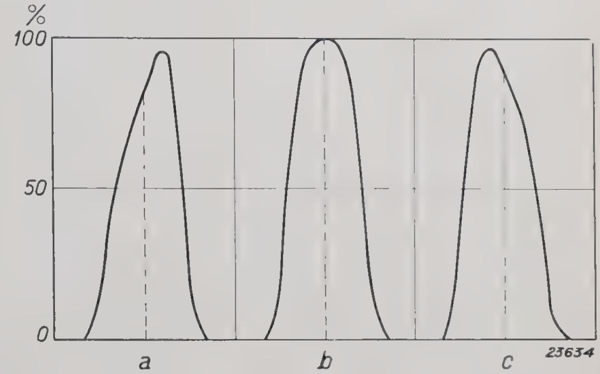

Fig. 8. The light transmission of the combination "Photoflux" (fig. 6) with shutter $^1/_{100}$ sec and F . 3.5 (fig. 5) when the half time value of the maximum opening of the shutter (given by vertical broken line) coincides with the peak value of the light emission of the lamp, and for the case when this falls 0.005 sec earlier (a) or later (c).

THE MOTION OF AN ELECTRON IN TWO-DIMENSIONAL ELECTROSTATIC FIELDS

by P. H. J. A. KLEYNEN.

Summary. The motion of a charged particle in a two-dimensional electrostatic field may be studied with the help of a model consisting of a stretched rubber membrane, over which a ball rolls.
The membrane is stretched in such a way that the tension of the surface is constant over the whole area, and the limiting conditions of the electrostatic field are applied to the model in the sense that the deviation of the membrane in the vertical direction is considered proportional to the electrostatic potential.
In several cases is studied the approximation of the path of a ball rolling over this rubber surface to the path of the electron in corresponding potential fields.

Introduction

The development of radio valve technique often demands a more or less accurate knowledge of the motion of the electrons in the electrostatic field of the valve electrodes. The mathematical investigation of this motion according to the ordinary laws of mechanics is impossible in most cases. The calculation of the potential field from the fairly arbitrary limiting conditions as they occur in practice is prevented as a rule by mathematical difficulties. This calculation is however only the first step in the right direction. If in some way or other the potential field has been successfully calculated, then the motion of the electron in this field must still be calculated with the aid of mechanics, which is in most cases no less difficult. It is obvious that other methods have been sought which lead more rapidly to a final result, although they may be less accurate than the mathematical analysis. In the first place one may try to determine the potential with the aid of probe measurements, or by making the lines of force visible. In addition one may look for other natural phenomena which may be described in the same way as the electric potential field by the differential equation of Laplace and which are often more amenable to measurement than the electrostatic potential. The most important method arising from these attempts is that of the electrolytic trough [1]). If the field is once known the equipotential lines may be drawn and the electron paths may be fairly easily constructed in this field of equipotential lines. Apparatus has also been developed recently which when coupled to an electrolytic trough, immediately draws the electron paths [2]). A disadvantage of all these methods is the unknown connection between the limiting conditions of the potential field and the final electron paths. If one desires to know the influence of the potential of one of the electrodes on the path, it is necessary to determine the field at various values of the potential, and then construct the path.

A method without these disadvantages is based upon the following principle [3]). If one imagines a membrane, in which only surface tensions can act, stretched horizontally in such a way that the surface tension is constant over the whole surface, and if the membrane is given a vertical deviation at several points, then the height of every point of the membrane when the deviations are small satisfies the differential equations of Laplace for two dimensions. If we give those points of the membrane which correspond to the electrodes of the valve a height which is proportional to the potential of those electrodes, then the height of every point is proportional to the potential in the electrostatic field. If we now allow a point mass to slide over the membrane without friction, with initial conditions similar to those of the electron in the potential field, then the horizontal projection of the motion of the point is similar in form to that of the electron, at least when the deviations are small. The deviation is considered small, when the slope of the membrane is small with respect to unity. Since

1) This method is based on the fact that the potential distribution is not disturbed when all the electrodes are immersed in an electrolyte. Because current flows between the electrodes, it is possible to measure the potential at any point with the aid of a probe electrode.

2) See for example D. Gabor, Nature **139**, 373 (1937) and D. B. Langmuir, Nature **139**, 1066 (1937).

3) See also W. R. Smythe, L. H. Rumbaugh and S. S. West, Phys. Rev. **45**, 724 (1934).

the membrane only approximately satisfies the two-dimensional Laplace differential equation, it is clear that all potential fields cannot be investigated by means of the above described model.

In the following we shall attempt to obtain an idea of the accuracy of this method by comparing several calculated electron paths with measurements taken from the model.

The differential equation of the membrane

A membrane in which only a surface tension acts will adjust itself upon deformation in such a way that the potential energy is at a minimum. Since this is proportional to the surface area, the membrane will assume a form which represents the smallest area under the given limiting conditions. Differential calculus gives the following equation for such a surface:

$$\left[1+\left(\frac{\delta h}{\delta y}\right)^2\right]\frac{\delta^2 h}{\delta x^2}-2\,\frac{\delta h}{\delta x}\,\frac{\delta h}{\delta y}\,\frac{\delta^2 h}{\delta x\,\delta y}+\left[1+\left(\frac{\delta h}{\delta x}\right)^2\right]\frac{\delta^2 h}{\delta y^2}=0.$$

h is here the deviation of a point of the membrane from the horizontal, x and y are rectangular coordinates in the plane of the undeformed membrane. The first derivatives occur here in the second power, and if they are small we may neglect their squares with respect to unity.

The equation then becomes:

$$\frac{\delta^2 h}{\delta x^2}+\frac{\delta^2 h}{\delta y^2}=0$$

and thus becomes identical with the Laplace equation.

We can also derive this equation from the equilibrium conditions of an element $dx\,dy$ of the surface.

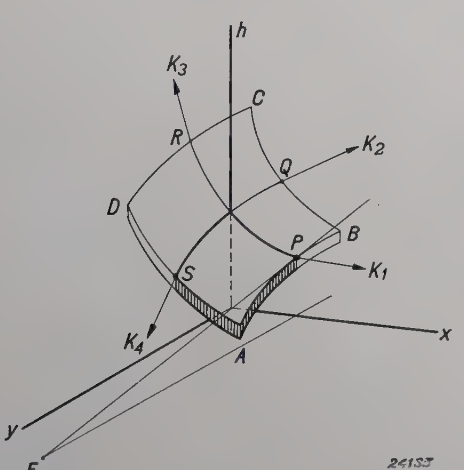

Fig. 1. Derivation of Laplace's differential equation from the equilibrium conditions of an element of the membrane.

In *fig. 1* $A\,B\,C\,D$ represents the surface element $dx\,dy$ which is formed by the intersection of the membrane by planes parallel with the coordinate planes. Forces act upon the intersecting planes, which touch the surface and are perpendicular to the curve of intersection. The element is in equilibrium, therefore the sum of these forces is zero:

$$K_1+K_2+K_3+K_4=0.$$

Of these forces we shall examine K_1. This force is perpendicular to the line of contact $P\,E$ on the intersection $A\,B$, and lies in the surface. If now the surface has only a slight slope from the xy plane, so that $P\,E$ makes only a small angle with the y-axis, then the forces K_1 and K_3 will lie practically parallel to the line $P\,R$ of the surface, and the forces K_2 and K_4 practically parallel to the line $A\,S$. In this case for the sum K_1+K_3 we find:

$$K_1+K_3=\sigma\,dy\left[\left(\frac{\delta h}{\delta x}\right)_P-\left(\frac{\delta h}{\delta x}\right)_R\right]=\sigma\,dy\,\frac{\delta^2 h}{\delta x^2}\,dx,$$

in which σ is the surface tension,

and for K_2+K_4

$$K_2+K_4=\sigma\,dx\,\frac{\delta^2 h}{\delta y^2}\,dy\,.$$

From the condition that the sum of all these forces must be equal to zero, it follows that:

$$\frac{\delta^2 h}{\delta x^2}+\frac{\delta^2 h}{\delta y^2}=0\,.$$

Thus we see that the membrane satisfies Laplace's differential equation if the slope is sufficiently small at each point.

If we now apply to the membrane and the potential field similar limiting conditions, i.e., if we give the membrane at points corresponding with the electrodes a height proportional to the potential of these electrodes, then at every point of the membrane the height is proportional to the potential at the corresponding points of the electrostatic field.

The analogy between the model and the electrostatic field can be worked out somewhat farther. It is clear that lines of equal height correspond to the equipotential lines of the field. The maximum slope at one point of the membrane is proportional to the field strength. In electrostatics furthermore, the charge on an electrode is proportional to the integral of the field strength along the edge of the electrode, and we may therefore determine the charge with the help of the model by measuring

the slope along the corresponding edge and integrating. Considering that the vertical component of the force acting on a linear element of the edge is proportional to the local slope of the membrane, this integral corresponds to the pressure of the membrane on the electrode.

The paths of the electron and the sliding point-mass

A consideration of the horizontally stretched membrane with the deviations applied, naturally gives rise to the question as to how far the horizontal projection of the motion of a point-mass, sliding over the surface without friction in the field of gravity, corresponds to that of the electron in the corresponding electrostatic case.

For the equation of motion of the electron we find

$$m\frac{\mathrm{d}^2x}{\mathrm{d}t^2} = -e\frac{\delta V}{\delta x}, \quad m\frac{\mathrm{d}^2y}{\mathrm{d}t^2} = -e\frac{\delta V}{\delta y} \quad \cdot \cdot (1)$$

We shall try to discover whether the differential equations for the projection of the motion of the point of mass on the horizontal plane have the same form.

Fig. 2 gives a vertical cross section through

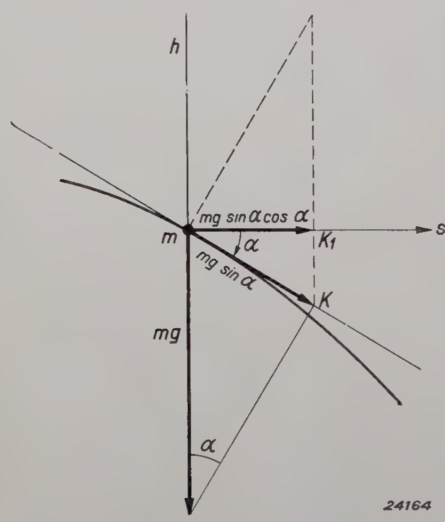

Fig. 2. Derivation of the forces acting on a point mass which slides on a sloping membrane without friction.

the membrane in the direction of greatest slope, with the forces which act on the point of mass m at a definite point of the membrane as a result of the force of gravity. It may be seen from the figure that the horizontal component K_1 of the resultant force K is equal to:

$$K_1 = m\,g\,\sin\alpha\cos\alpha,$$

where α is the angle of maximum slope.

Furthermore:

$$\tan\alpha = -\frac{\mathrm{d}\,h}{\mathrm{d}\,s},$$

therefore:

$$K_1 = -mg\cdot\frac{\dfrac{\mathrm{d}\,h}{\mathrm{d}\,s}}{1+\left(\dfrac{\mathrm{d}\,h}{\mathrm{d}\,s}\right)^2}.$$

The components of the force in the x and y directions respectively are then:

$$\left.\begin{aligned} K_x &= m\frac{\mathrm{d}^2x}{\mathrm{d}\,t^2} = -mg\frac{\dfrac{\delta\,h}{\delta\,x}}{1+\left(\dfrac{\mathrm{d}\,h}{\mathrm{d}\,s}\right)^2},\\ K_y &= m\frac{\mathrm{d}^2y}{\mathrm{d}\,t^2} = -mg\frac{\dfrac{\delta\,h}{\delta\,y}}{1+\left(\dfrac{\mathrm{d}\,h}{\mathrm{d}\,s}\right)^2}. \end{aligned}\right\} \quad (2)$$

If we now again assume that the slope of the membrane is small, we may neglect $\left(\frac{\mathrm{d}h}{\mathrm{d}s}\right)^2$ with respect to unity and equation (2) becomes:

$$m\frac{\mathrm{d}^2x}{\mathrm{d}\,t^2} = -mg\frac{\delta\,h}{\delta\,x}, \qquad m\frac{\mathrm{d}^2y}{\mathrm{d}\,t^2} = -mg\frac{\delta\,h}{\delta\,y} \quad \cdot (3)$$

These equations are indeed equivalent to (1) and with the proper choice of the boundary conditions the motions of the electron and of the point-mass will therefore also be equivalent.

Construction and testing of the membrane

Rubber is practically the only suitable material for the membrane. A sheet of rubber about 1 m in diameter, upon which rectangular and polar coordinates have been drawn before stretching, is stretched in a metal ring in such a way that the tension is constant over the whole area. If the coordinates drawn on the sheet are uniformly enlarged by the stretching, it may be assumed that the surface tension is everywhere constant. *Fig. 3* gives an idea of the arrangement. The preliminary check tests were carried out with a sheet 1 mm thick which was fastened with a stretch of 3 per cent. It was found that the irregularities in the tension at the edge which result from the method of fastening, extend no farther than about 5 cm toward the centre.

The sheet is set up horizontally on a table with a horizontal and plane surface. The plane projection

of the electrodes of the electrostatic arrangement is now applied to the sheet on a suitable scale and at the positions of the electrodes the sheet is given deviations proportional to the potential of these electrodes by means of blocks or rings placed on or beneath the sheet. Since the charge of the electron is negative we plot the positive potential in a downward direction and the negative in an upward direction.

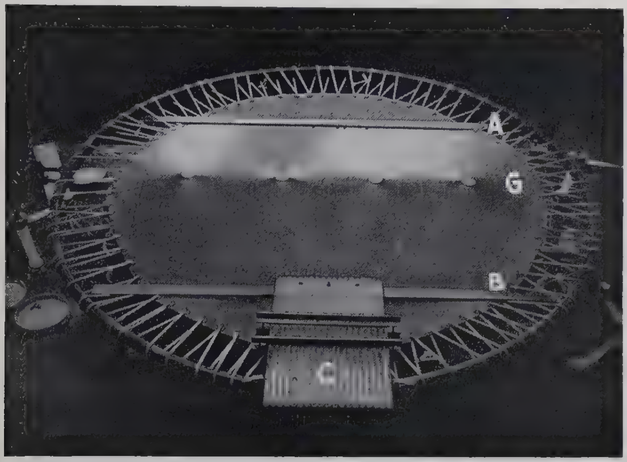

Fig. 3. Survey of the arrangement of the model. The rubber membrane is uniformly stretched in a horizontal iron ring. The height is fixed by two horizontal strips A and B. and by four cylinders G placed under the membrane.

From the form of the differential equations of the membrane and from the motion of the point-mass it follows that the path is independent of the scale on which the model is built. This holds not only for the linear dimensions but also for the ratio in which potential is transformed into height. It is known that in the electrical case also the shape of the path of the electrons does not change when the linear dimensions of the electrode system or the potential are increased proportionally.

Fig. 3 shows the potential field of a negative grid between two parallel positive plates. We shall return to discuss this in more detail later. The circles at G represent the intersections of the grid wires with a plane perpendicular to the grid wires. The straight lines at A and B are the lines where this plane cuts the parallel plates. The grid G consists in this case of four equally high cylinders which are placed on the table under the membrane. The field between the two middle grid supports is practically unchanged when the number of cylinders is increased to the left and right.

After the membrane is set up with the correct boundary conditions, the problem is to have a point-mass slide over the surface without friction. It is found, however, that with a membrane of rubber the coëfficient of sliding friction is much too large to produce a usable result. Therefore a rolling ball is substituted for the sliding point-mass, keeping in mind that the rotational energy of the ball and the rolling friction may be the causes of deviations in the path of the ball from that of a point mass sliding without friction. When we attempt to obtain an idea of this source of error in a mathematical way, we meet with great difficulties. We may introduce the moment of inertia and also the rolling friction into the differential equations of the motion in some form or other, but it is difficult to draw a conclusion from the form of these equations about the influence of these factors on the form of the path.

The only possibility then remaining is an attempt to show the errors in an experimental way. This can be done in different ways. In the first place with a given arrangement of the membrane we may change the friction and the moment of inertia of the ball, and then compare the paths of the ball. The friction for example may be changed by spreading a thin layer of powdered talcum on the membrane and the moment of inertia by taking balls with a heavy core and a light covering or the reverse. A simple calculation shows that a ball can be made with a platinum or a tungsten core and a aluminium covering, which has the same mass and the same diameter as a steel ball, but whose moment of inertia is one third as large. A second method consists in using an arrangement in which the motion of a sliding point mass is theoretically accurately known, and comparing this with the motion of the ball. A third method consists in the comparison of the path of the ball with the path of electrons, the shape of whose path is already completely known from mathematical or experimental investigation.

We have applied the second method to the motion of a ball on a flat sloping plane, and the third method to the motion of an electron in the field of a slit between two parallel plates.

The motion of a ball on a flat sloping plane

If the coördinates are placed as in *fig. 4*, the path of a point mass sliding without friction over

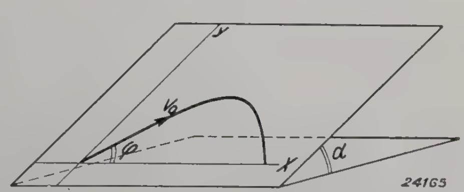

Fig. 4. Motion of a point mass on a sloping plane.

a flat sloping plane is a parabola with the equation:

$$y = x \tan \varphi - \frac{g \sin a}{2\, v_0^2 \cos^2\varphi}\, x^2, \quad . \quad . \quad (4)$$

where v_0 is the initial velocity and g the acceleration due to gravity.

The motion of the ball on the sloping plane may be recorded as follows. The highly polished ball is allowed to leave the source and is photographed during its motion while the arrangement is illuminated with a high pressure mercury lamp burning on single-phase alternating current. Since this lamp is periodically lighted and extinguished, and the period is equal to half that of the alternating current, the path of the ball is visible as a dotted line (see *fig. 5*). If the lamp is running on

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Fig. 5. Photograph of the path of a polished ball on flat sloping plane. The ball is illuminated by a lamp which is periodically extinguished, so that the track is visible as a dotted line.

50 period alternating current, the distance between two adjacent dots is covered in one-hundredth of a second, and the speed of the ball can thus be calculated directly from this distance. The angle φ at which the ball leaves the source and the slope of the membrane can be measured, and with the aid of (4) we can calculate the path which is followed by a point mass sliding without friction when it leaves the source under the same conditions. It must in addition be noted that the rotational energy of a truly rolling homogeneous sphere is $^2/_7$ of the total kinetic energy, so that we must multiply the velocity derived from fig. 5 by $\sqrt{7/5}$ in order to obtain the corresponding velocity of the sliding point mass.

If we carry out the calculation in this case and compare the parabola with the path of the ball of fig. 5 we find that the top of the parabola lies about 1 per cent higher than the top of the path of the ball, while the distance between the points at which the parabola cuts the x-axis is about 7 per cent greater than that between the corresponding points of the path of the ball. If the friction is increased slightly this distance decreases sharply, so that we have in this fact a good criterion for the influence of friction.

The electron in the field of a slit.

The foregoing case gives us an idea of the influence on the path of the friction alone. In the case now to be examined, the motion will be affected by sources of error which are connected with the fact that the slope of the membrane is not infinitesimally small, in addition to friction.

We imagine a plane plate S with a straight slit parallel to and situated at equal distances from two parallel plates A and K. *Fig. 6* shows a cross sec-

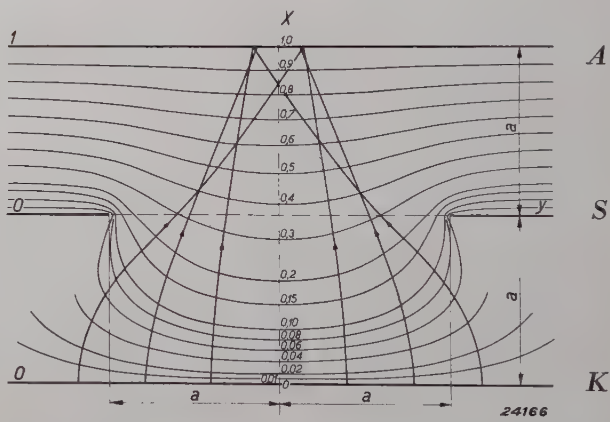

Fig. 6. The potential field of a slit between two parallel plates with the electron paths constructed in this field.

tion perpendicular to the plates and to the slit. With the aid of the theory of the function of a complex variable the potential field of such an electrode arrangement may be calculated. If we apply to plate A a potential equal to unity and to the slit and the other plate a potential zero, and if we choose the width of the slit and the distances between the plates as given in fig. 6, the potential V is given by the following:

$$V = \frac{x}{2a} + \frac{1}{2\pi} \operatorname{arc\ cos} \left[\frac{\cosh^2 \frac{\pi y}{2a} - \sin^2 \frac{\pi x}{2a} - \sqrt{\left(6{,}295 + \cosh^2 \frac{\pi y}{2a} - \sin^2 \frac{\pi x}{2a}\right)^2 - 25{,}180 \cosh^2 \frac{\pi y}{2a} \cos^2 \frac{\pi x}{2a}}}{6{,}295} \right].$$

The calculation of V from this formula gives the equipotential lines shown in fig. 6, and in this field of equipotential lines we can construct the electron path in a simple manner. The construction is based on the same principle as the apparatus which derives the shape of the path from a potential and gradient measurement in the electrolytic trough (see for example D. Gabor and D. B. Langmuir, loc. cit.). The radius of curvature of the path can be calculated at every point from the potential and the gradient perpendicular to the path. If we now allow the electrons to leave with a velocity zero at the plate K, then the construction gives the paths reproduced in fig. 6.

In the rubber model the plates K and S are at equal heights, while A is somewhat lower. At K we allow 7 balls to roll away rapidly one after another with no initial velocity. We then obtain the paths as shown in *fig.* 7.

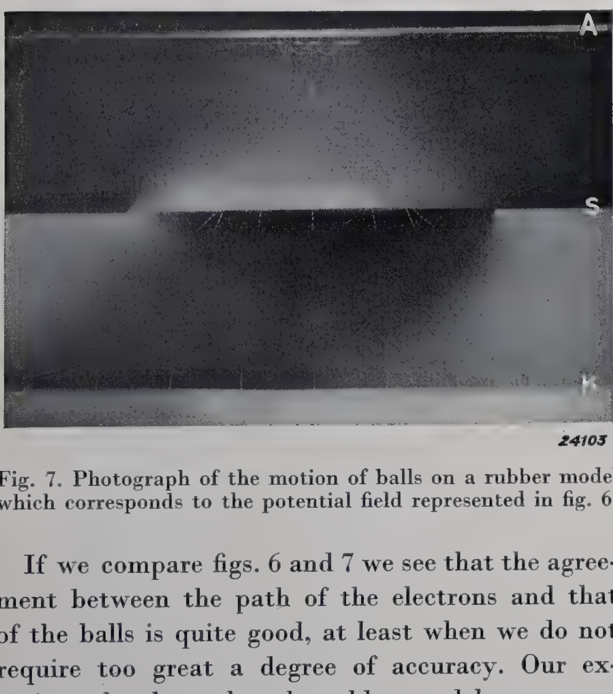

Fig. 7. Photograph of the motion of balls on a rubber model which corresponds to the potential field represented in fig. 6.

If we compare figs. 6 and 7 we see that the agreement between the path of the electrons and that of the balls is quite good, at least when we do not require too great a degree of accuracy. Our experience has been that the rubber model may prove very valuable for design purposes in radio valve technology. In other cases, however, such as for example the determination of distortions of the image in electron optics, one must be careful, and only more elaborate tests of the above mentioned type will be able to give assurance that the accuracy attained is sufficient.

Limitations

As we have already mentioned, the application is limited to two-dimensional cases, or to three-dimensional potential fields in which the potential is independent of one of the coordinates. In general therefore rotation-symmetrical fields for example cannot be investigated by means of the model.

With greater electron densities we must moreover take into account the influence of the space charge on the potential field. The difficulty here is that the change of potential due to this space charge is dependent on the very electron path which we want to determine. If however the space charge is not too great, one can first determine the paths on the assumption that the space charge is zero, and with the help of these paths one may then make an estimation of the space charge. The lowering of the potential which results from this may then be applied to the model by increasing the height of the membrane locally proportional to this lowering of potential, and the path may then be determined once more. If this process is repeated the results usually converge quite rapidly to the correct solution.

Several other applications

The rubber model may also be used to measure capacities of two-dimensional electrode arrangements. If, for example, it is desired to determine the capacity of an electrode with respect to another set of electrodes, in the electrostatic case that electrode may be given a potential difference V with respect to the other electrodes which all have the same potential, and then the charge of the electrode under consideration may be determined. This charge is, per cm axial length, equal to $\frac{1}{4\pi}$ times the integral of the field strength along the edge of the electrode. The ratio between charge and potential difference then gives the capacity per cm. In the rubber model the potential difference V is equivalent to a difference in height h between the electrode named and the others, while instead of the field strength we integrate the slope of the membrane along the edge of the electrode. We then find the following for the capacity:

$$c_1 = \frac{1}{4\pi h}\int_S \tan\alpha\,\mathrm{d}s = \frac{1}{4\pi}\cdot\frac{D}{\sigma h}, \text{ where}$$

c_1 = capacity per cm axial length
h = difference in height between the electrodes
α = slope of the membrane at the edge of the electrode
S = limit of the electrode
D = pressure on the electrode
σ = tension of the rubber membrane.

The dimensions of the electrode system may be applied to the model in any arbitrary enlargement,

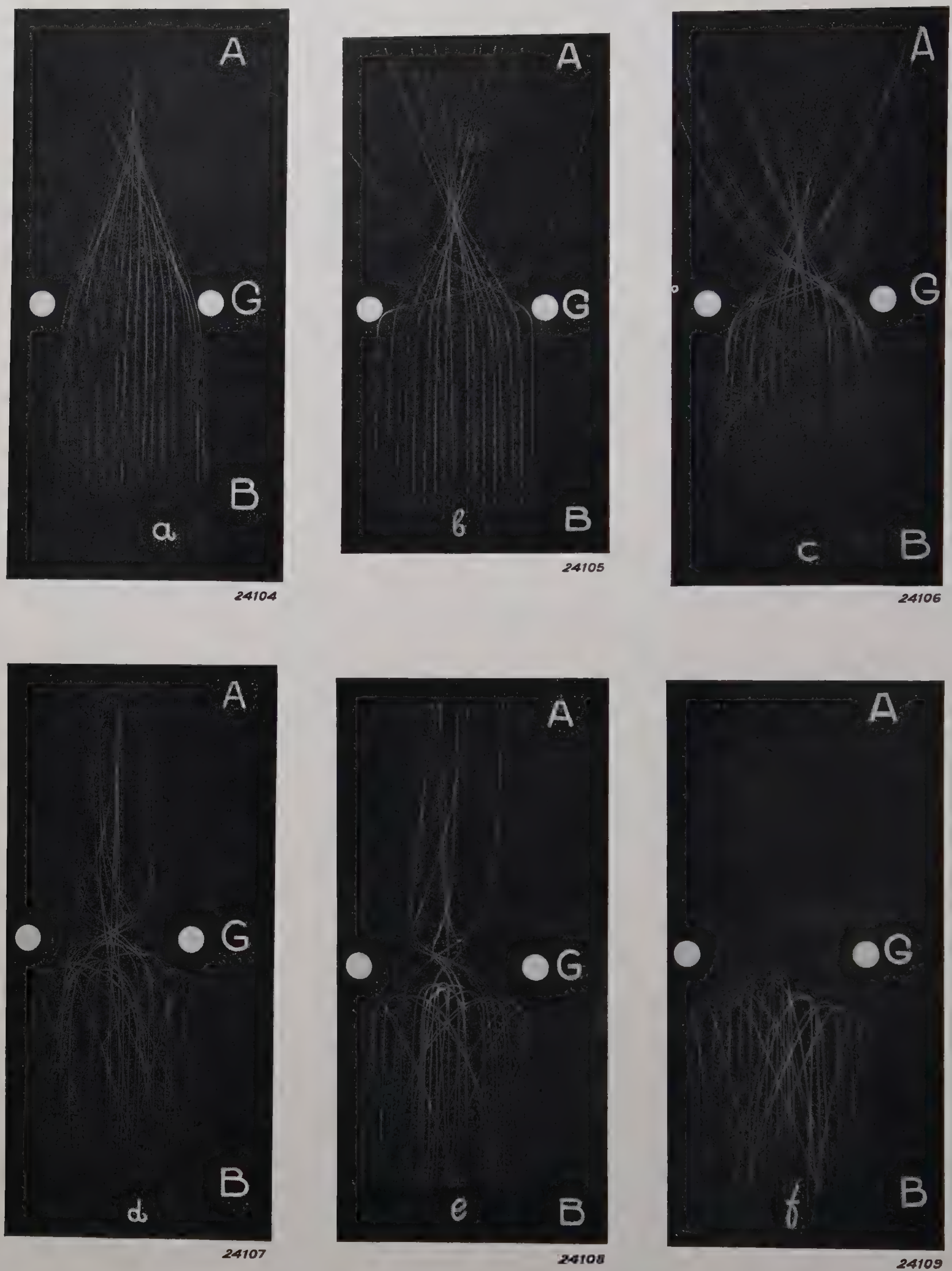

Fig. 8 a - f. Photographs of paths of balls, obtained by shooting a parallel "beam" of balls (in the arrangement of fig. 3) with a constant velocity perpendicular to the line joining the grid supports. The different figures are obtained by taking the height of *G* successively a little greater.

since the capacity of a two-dimensional arrangement per cm axial length is without dimensions. The slope may be measured by laying a small mirror on the membrane, and then with the aid of a telescope with a divided scale and a level recording the angle from the normal.

A photograph of the path of the ball further enables us to determine the time of transition of the electron in the corresponding electrostatic case. We have already seen that we may calculate the velocity of the ball at every point from the distance between two dots. This velocity is, however, dependent upon the dimensions of the model and also is much smaller than that of the electron. Since, however, the motions of the electron and of the point mass sliding without friction are similar in shape, we may consider it reasonable that there is a constant relation between the time elements of the motions of the electron and of the point. This relation may be determined from the differential equations (1) and (3), which for this purpose may be written in the following form:

$$\frac{d^2 x_e}{d t_e^2} = -\frac{e}{m}\frac{\delta V}{\delta x_e} \cdot \text{(1a)} \quad , \quad \frac{d^2 x_m}{d t_m^2} = -g\frac{\delta h}{\delta x_m} \,. \text{(3a)}$$

$$\frac{d^2 y_e}{d t_e^2} = -\frac{e}{m}\frac{\delta V}{\delta y_e} \cdot \text{(1b)} \quad , \quad \frac{d^2 y_m}{d t_m^2} = -g\frac{\delta h}{\delta y_m} \,. \text{(3b)}$$

The subscripts e and m indicate that the quantities concerned are taken from the electrostatic case and from the rubber model respectively. Since the ratio between the size elements (dx_e/dx_m and dy_e/dy_m) is known, and since we also know to what difference in height in the model the unit tension corresponds, we may determine from the equation the ratio of the time elements. We may for example divide equations (1a) and (3a) by each other, and then obtain from the resulting equation for the ratio of the time elements:

$$\frac{dt_m}{dt_e} = \frac{x_m}{x_e}\sqrt{\frac{e\,V}{g\,m\,h}} \,.$$

If we assume that the ball is homogeneous, and that it rolls without slipping, then, because of its rotational energy, its velocity at every point is at the most $\sqrt{5/7}$ times that of the sliding point mass under the same conditions, and in the case of the motion of a ball on a flat plane this factor is exactly $\sqrt{5/7}$. We shall be able to account for the moment of inertia to a large degree by choosing the time elements with the ball motion a factor $\sqrt{7/5}$ greater. We then find for the ratio:

$$\frac{dt_m}{dt_e} = \frac{x_m}{x_e}\sqrt{\frac{7e\,V}{5mgh}} \,.$$

Finally we shall give an example of the application to the subject of radio valves. If we imagine a positive anode and in front of it a grid at a lower potential, we may then ask how the electrons of a parallel beam, which are sent out with a definite velocity perpendicular to the anode, are deflected by the wires of the grid. The model then has the form shown in fig. 3. In this figure A and G represent the anode and the grid respectively. The balls are freed perpendicularly to B with a speed determined by the slope of a set of tubes C, in which the balls have attained their speed under the influence of gravity. The potential of plane B is determined at the same time by this speed, and with it those of A and G.

Fig. 8 gives a picture of the paths of the balls with various heights of G, that is with various potentials of the grid. In the first recording (fig. 8a) the potential of the grid was such that all the electrons were able to pass the grid, and we see how the electrons come together in a focus point between grid and anode. If we lower the potential of the grid, the focal distance becomes shorter and the electrons at the edge of the beam turn back. In fig. 8f the grid is so negative that all the electrons are reflected.

An important advantage of the above-described method is, that with very little trouble one may investigate the influence of a change of arrangement or potential of the electrodes. It is only necessary to shift the electrodes in the model slightly or to increase their height, and then to observe the paths of the balls.

A VIBRATOR FOR THE CONNECTION OF ALTERNATING CURRENT RECEIVING SETS TO THE DIRECT CURRENT MAINS

by J. W. ALEXANDER.

Summary. Receiving sets can be more advantage ously constructed for alternating current than for direct current supply. In this article a vibrator is described which transforms direct voltage into alternating voltage. By introducing this instrument into the circuit, an alternating current receiver is rendered suitable for supply with direct current of the same mains voltage without further changes in its construction.

Introduction

In the case of radio sets which are fed from alternating current mains it is very simple to obtain the necessary heating voltage by means of a transformer for the valves connected in parallel. The necessary anode voltage is delivered *via* another winding on the same transformer and a rectifier valve. By giving the primary of the transformer several tappings the set may be made suitable for the ordinary voltages of 100-155 volts and 200-250 volts.

If, however, the apparatus is to be supplied from direct current mains, several difficulties arise which are all connected with the fact that no transformer can be used to adapt the apparatus to the voltage available. It is for instance necessary to connect the filaments of the different valves in series instead of in parallel, which introduces difficulties, since it is not easy with this method of connection to avoid the presence of hum from the mains. The provision of sufficient insulation to avoid the danger of accidental contact is much more difficult with direct current sets than with alternating current sets. Moreover it is practically impossible to insure that the set works equally well on all the mains voltages existing between 110 and 250 volts direct current.

The above-mentioned difficulties have led to an attempt to remove the disadvantage of direct current supply by transforming the direct voltage into alternating voltage by means of a vibrator. By connecting this instrument in circuit with the set, an alternating current receiver is rendered suitable for supply with direct current of the same range of mains voltages without the necessity of further alterations in its structure.

Principle and functioning of the vibrator

The circuit of the vibrator is reproduced diagrammatically in *fig. 1*. The vibrator may be considered as a double commutator, consisting of the springs A_1 and A_2, electrically separated, but mechanically connected, which can make contact alternately with K_{11} and K_{12}, and with K_{22} and K_{21} respectively. The springs are moved by an electromagnet M acting on an armature, which is mechanically

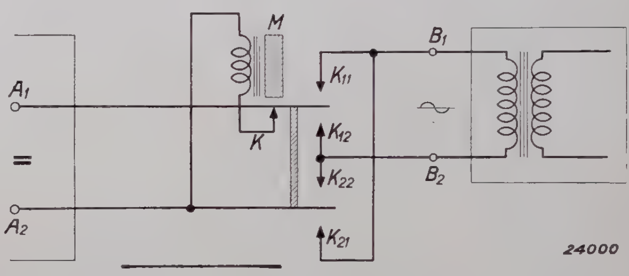

Fig. 1. Circuit diagram of the vibrator.

connected with the springs A_1 and A_2. If the vibrator is at rest and the terminals A_1 and A_2 are not yet connected to the mains then spring A_1 is connected to the coil of the magnet by means of the contact K. When the vibrator is switched on the electromagnet will attract the armature, thereby making contact between A_1 and B_1 (*via* K_{11}) and between A_2 and B_2 (*via* K_{22}). The contact K is broken. The springs bend farther to their maximum deviation, return and, owing to their inertia, pass the zero point to make contact between A_1 and B_2 *via* K_{12}, and between A_2 and B_1 *via* K_{21}. Upon passing the point of rest the magnet is again brought into circuit, etc. In this way the mechanism continues to function.

It is important to note, that an essential condition for the functioning of this mechanism is that growth of current in the magnet be retarded by its own self-induction.

When the contact K is closed and the armature is moving from the point of rest to the maximum deviation, the armature is retarded; when the armature is moving from the maximum deviation toward the position of rest, it is accelerated. If the self-induction of the electromagnet had no retarding

effect on the current, the retarding force would be equal to the accelerating force, and the total energy applied to the armature would be zero. As a consequence of the self-induction an excess of applied energy occurs, and this is necessary in order to keep the mechanism working.

The result of making contact alternately at K_{11} and K_{22}, and K_{12} and K_{21} respectively is that the transformer is given a voltage of alternating sign, which, apart from the transformation ratio, induces the same voltage on the secondary side. If the transformer is shunted by a resistance, the secondary current has the form of *fig. 2b*, while a primary current flows as shown in *fig. 2a*, an interrupted direct current.

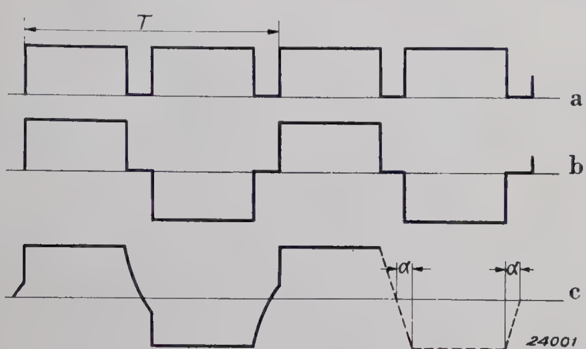

Fig. 2. a). Intermittent direct current taken from the input side of the vibrator.
b). Secondary voltage which would be delivered by a vibrator according to the circuit of fig. 1. T is the time of one period.
c). Alternating voltage delivered when the contacts are bridged with the help of condensers. The broken lines indicate how the curve may be idealized.

A vibrator constructed on this principle would have the disadvantage that at each half period upon commutation a high voltage would occur at the contacts where the circuit is broken, as a result of the self-induction of the transformer. This voltage surge can, however, be avoided by shunting the contacts with condensers. This is indicated in *fig. 3*. We assume that the vibrating spring is just on the point of leaving the contacts K_{11} and K_{22}.

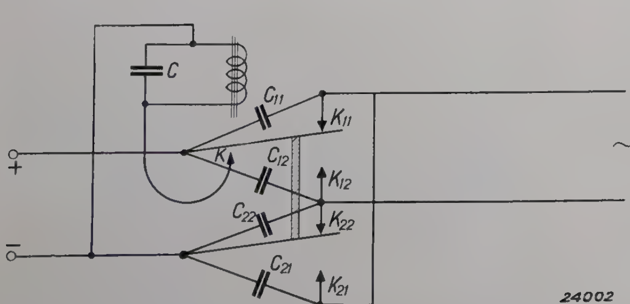

Fig. 3. By bridging the contacts with condensers high voltage commutation surges can be avoided every half period.

The current from the mains will then immediately be interrupted. The current through the transformer, however, will be able to continue and charge the condensers C_{11} and C_{22}, and will decay in a manner determined chiefly by the magnitude of the load impedance between B_1 and B_2, by the capacity of the condensers and the self-induction of the transformer. The slope $\mathrm{d}i/\mathrm{d}t$ of the current-time curve thus remains finite, and consequently also the output potential.

When the vibrating spring reaches the opposite contacts K_{12} and K_{21} the residual charge of the condensers C_{12} and C_{21} is short circuited. A strong short-circuit current surge is thereby caused, and the voltage at the output terminals immediately assumes the value of the mains voltage. In fig. 2*c* the diagram of the output potential at B_1, B_2 is given.

The disadvantage of the above-mentioned strong short-circuit current surge may be easily overcome by including in the circuit, in series with the condensers, a resistance which limits the short-circuit currents.

In addition to the parts sketched in fig. 1 the vibrator includes also two filters which protect the mains and the set respectively from high-frequency oscillations, which might be generated in the apparatus by the sudden current variations upon breaking the contacts. Reception entirely free of disturbances is by this means attained for frequencies of 150 kc/sec and higher, that is, over the whole broadcasting range.

Mechanical construction and mounting of the vibrator

Fig. 4 gives the mechanical construction. The vibrator consists of a frame R, which at the same time forms part of the electromagnet with the coil S. At a short distance from the core is the armature M fixed to and insulated from the two springs A_1 and A_2, to which the direct voltage is applied. These springs are insulated from the frame by means of mica plates P. Not far from the points where they are clamped the springs A_1 and A_2 have a u-shaped stiffened section, upon which the tungsten contacts K_{11}, K_{12}, K_{21} and K_{22} are fastened. The opposite contact plates are mounted a short distance away on side springs, which are insulated from the frame. Together with the u-shaped section a second spring is attached to spring A_2, with the contact K for connecting the current through the coil S of the magnet. When the contacts are open the middle springs A_1 and A_2 vibrate about the points where they are clamped;

during the greater part of their deviation, however, they are constrained by the side springs, so that the natural frequency of the middle springs

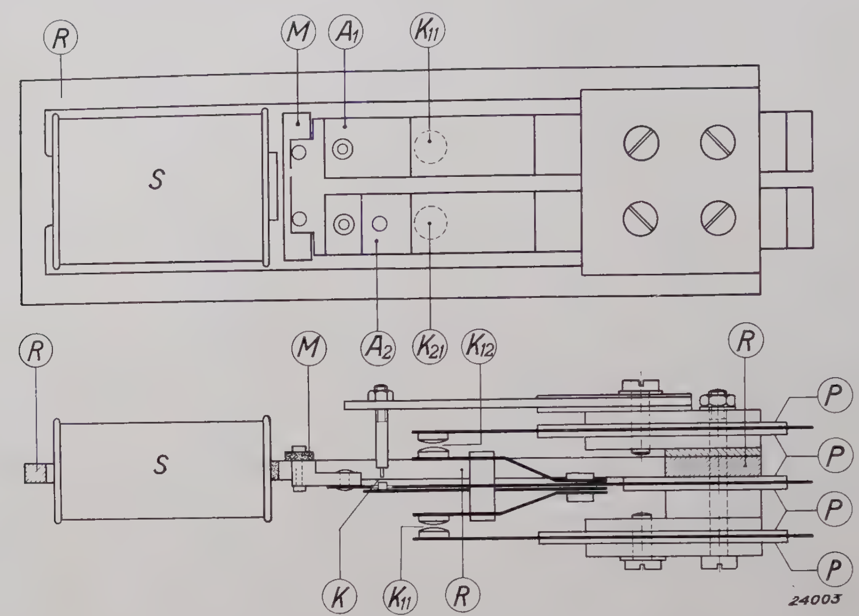

Fig. 4. Mechanical construction of the vibrator.

weighted with the armature is considerably increased. It is about 100 c/s.

The condensers and coils of the interference filter are placed in a separate can, which is connected to the vibrator by means of six contact bushes. Two of the six pins serve to lead in the direct current, two to lead out the alternating current and two for the interference condenser C (fig. 3), which bridges the magnet coil and which also is placed in the separate can. *fig. 5* shows the different parts separately; *fig. 6*

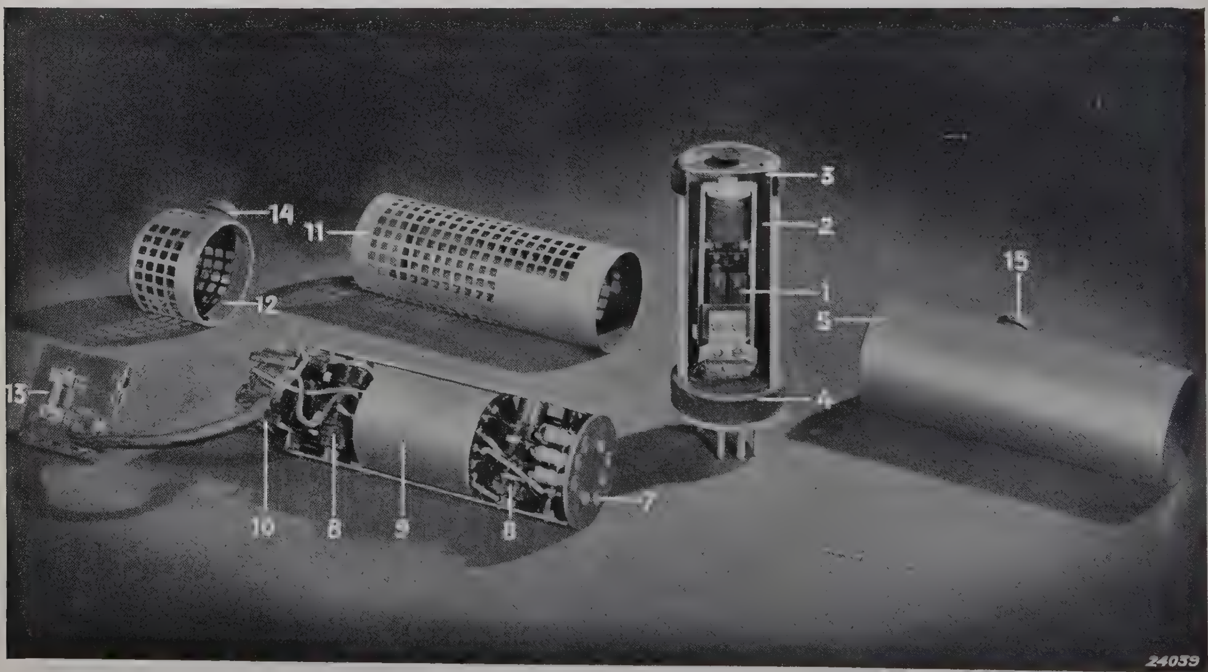

Fig. 5. Parts of the vibrator. The vibrating element *1* is hung on springs in the can *2* which is sealed by the rubber cover *4*, which also serves to close can *5* surrounding can *2*. The vibrator with its six pins is inserted into the contact bushes *7* of the interference filter component. *8* chokes and *9* condensers of the interference filter element; *10* coils with iron cores, *11* and *12* ventilation of the interference filter element. *13* fuse, *14* and *15* rubber studs for flexible and damped mounting in the receiving set.

shows the vibrator joined to the interference components.

In mounting the vibrator account must be taken of the fact that the alternating current generated is not sinusoidal, but contains higher harmonics, and chiefly odd harmonics of about 300, 500, 700 periods per second [1]). By careful layout of the receiver components, and, if necessary, by electric or magnetic shielding of the parts sensitive to those influences, the penetration of those harmonics into the receiving set can be avoided.

Another important point in mounting is the the can which contains the vibrator, air-tight, and surrounding it with a second can which is closed by a rubber seal through which the connecting wires are led. In this way chiefly vibrations with a higher frequency are damped. The vibrations with a lower frequency, which are propagated more through the solid material, are sufficiently arrested by mounting the vibrator itself in the inner can with the help of springs, so that it has some fredom of movement.

The value of the voltage delivered by the vibrator is closely related to the voltage of the direct

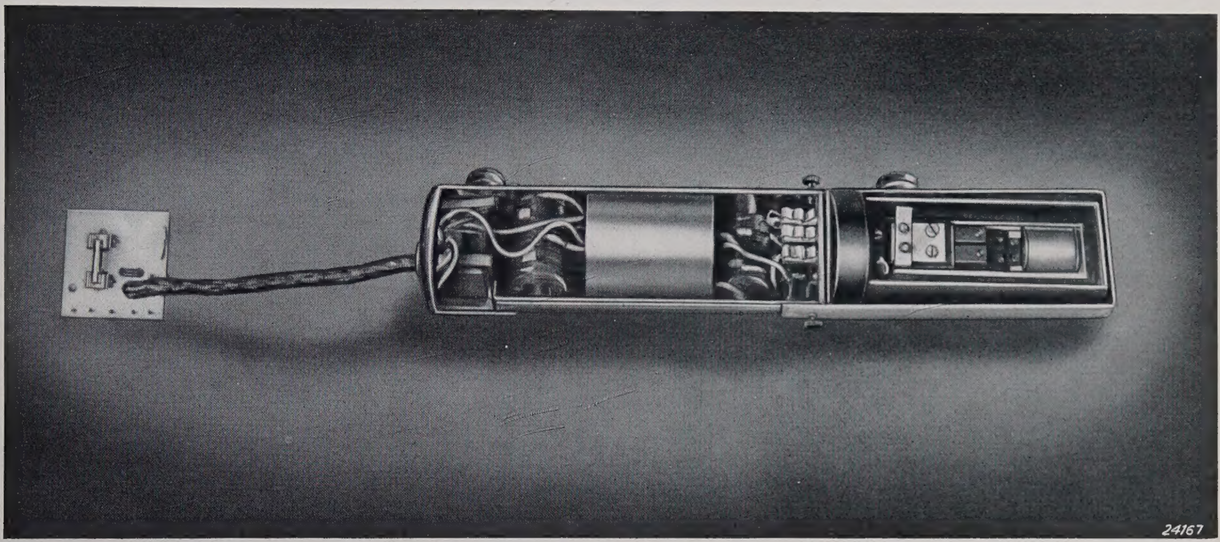

Fig. 6. Vibrator with non-interference components.

damping of the acoustic vibrations caused by the vibrator. These vibrations are conducted not only by the solid material, but also by the air. The solution of this problem has been found by making

current mains to which the vibrator is connected. If it is supplied with 220 volts, an alternating voltage of about 200 volts is obtained. This is clear when we remember that the maximum value of the alternating voltage delivered by the vibrator is equal to the direct voltage V_0. The effective value of the alternating voltage will therefore certainly be lower. If we represent the voltage schematically in the way used in fig. 2c, we find the following:

$$V_{\text{eff}} = V_0 \sqrt{1 - \frac{4\,a}{3\,\pi}}\,.$$

The effective voltage is thus lower than V_0. The output voltage of the vibrator is in that case not connected to the 220 volts terminals of the radio set, but to a special tap which is introduced for this purpose.

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[1]) If we idealize the output voltage of the vibrator in the manner represented in fig. 2c we obtain the following Fourier series:

$$V = \frac{4}{\pi} V_0 \left[\frac{\sin a}{a} \sin \frac{2\,\pi\,t}{T} + \frac{\sin 3\,a}{3\,a} \sin 3\,\frac{2\,\pi\,t}{T} + \right.$$
$$\left. + \frac{\sin 5\,a}{5\,a} \sin 5\,\frac{2\,\pi\,t}{T} + \cdot\,\cdot \right].$$

Thus odd harmonics occur. At the input end of the vibrator an intermittent direct current is introduced, which, as may be seen from fig. 2, contains a main frequency $2/T$ and harmonics of this, which are therefore even harmonics of $1/T$. These alternating currents might superimpose alternating voltages on the mains, and thereby cause disturbances in receiving sets in the neighbourhood. For this reason the alternating currents are prevented from entering the main by the introduction of a choke coil.

THE INVESTIGATION OF THE MACRO-STRUCTURE OF RAW MATERIALS AND PRODUCTS BY MEANS OF X-RAYS II.

by J. E. DE GRAAF.

Two kinds of information may be obtained from X-ray examination: information about the strength as it is decreased by defects, and information about the cause of these defects.

The first type of information can seldom be obtained by a photographic method like the absorption method. The shape and dimensions of a flaw and thus the decrease in cross section may, it is true, be indicated, but this is only of value when the weakening due to the notch effect is not greater than that corresponding to the decrease in cross-section. Only with spherical flaws such as gas bubbles would one be able to determine the weakening. In all other cases the evaluation of the strength requires wide experience if it is to be at all reliable to any degree.

The question of the diagnosis of the flaw, although it requires much experience, may usually be answered conclusively. Examples of this will be given here and in subsequent articles.

The discovery of flaws in riveted joints

Although riveting as a method of making joints hast lost much ground to welding, and particularly

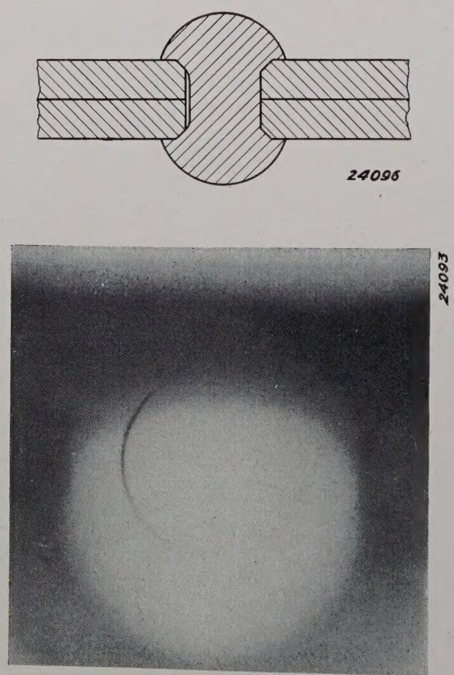

Fig. 1. Cross section and X-ray photograph of a rivet which does not entirely fill the holes.

to electric welding, it still remains an important working method. In certain cases it has retained a special position next to welding, for example, in the construction of parts of ships which are made of kinds of steel which easily crack upon non-uniform cooling. In such cases it is often undesirable to make the connections between parts welded in the shops by welding in the open air, while it is still quite possible to do this by riveting. As in welding, it is very important that the riveted joint be reliable, and not weakened by riveting flaws.

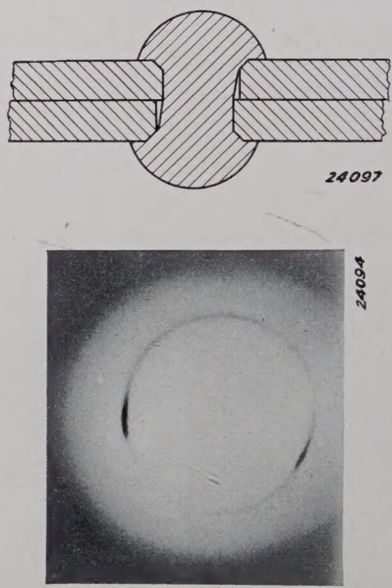

Fig. 2. The holes in the two plates of the riveted joint do not lie directly one above the other.

A riveting flaw which occurs fairly often is shown in *fig. 1:* the loose rivet. In this case, the rivet does not completely fill the hole, and the cylindrical fissure between the hole and the rivet is shown as a dark circle on the photographic negative. When the holes in the parts to be joined are not situated exactly one above the other (*fig. 2*), such circular arcs also will occur, but in this case, they do not coincide, but form parts of two non-concentric circles (for a true diagnosis it is of course necessary that the rays be incident perpendicularly to the plate). In the photographs of both of these flaws a sharply limited outer circle is seen which corresponds in diameter with that of the holes bored. By means of a circle drawn on transparent paper one may usually see clearly whether the flaw consists of arcs of one or of two circles. A third flaw which also gives rise to a circular image is shown in *fig. 3:* the head of the rivet was badly formed and

later worked over to give the correct external shape. This flaw, however, in contrast to the two foregoing ones, shows an unsharp outer limit in the photograph, while the average diameter is greater than

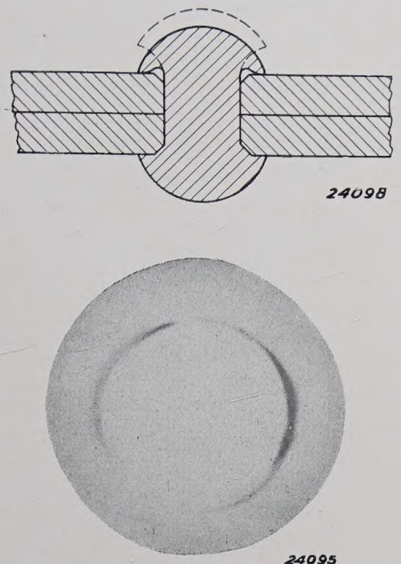

Fig. 3. The head of the rivet is badly formed and later worked over into the correct external shape.

the diameter of the rivet hole. This also can usually be detected with the circle drawn on transparent paper. Since the significance of these three flaws is not the same, the importance of their individual recognition will be clear.

The crooked head, which occurs quite often in hand riveting, is shown very clearly in an X-ray photograph as an irregular grey border, as may be seen in the upper left had corner of *fig. 4*.

In the boiler plate itself very dangerous cracks may occur, not only around the rivet holes (*fig. 4*), but between successive holes. Cracks are particularly important since they always have the tendency

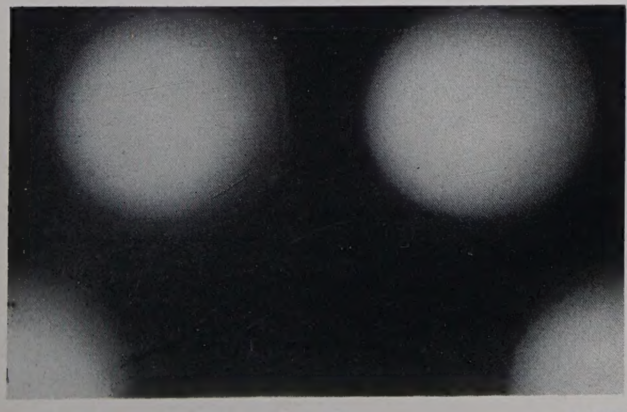

Fig. 4. X-ray photograph of cracks in a boiler plate in and around the rivet holes (cf. fig. 5a).

to become larger, which fact may be ascribed to the high concentrations of stress at their ends (notch effect). The detection of cracks is, however, not always simple. In the case of *fig. 5a*, in which the

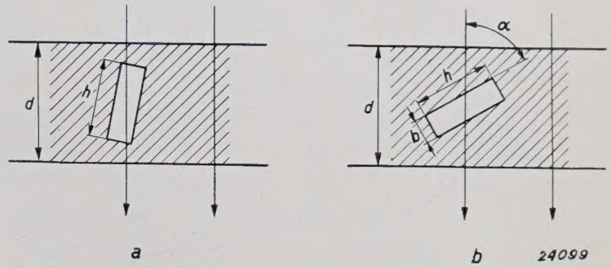

Fig. 5. The photography of cracks at different angles between crack and X-rays.

angle between the X-rays and the crack is small, the local decrease in thickness is equal to h, the whole height of the crack. In fig. 5*b* the decrease in thickness is only $b/\sin \alpha$. With increasing α therefore the difference in blackening caused by the crack will decrease rapidly. *Fig. 6* gives an

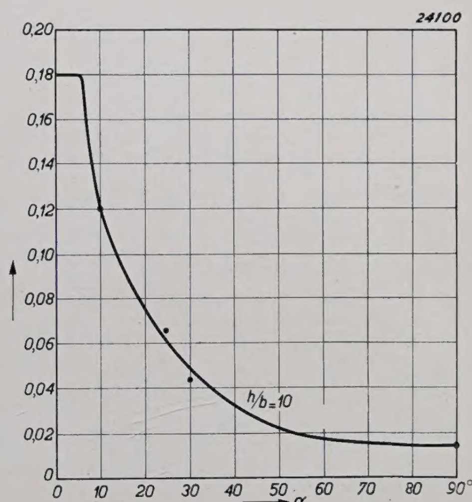

Fig. 6. Contrast between the picture of the crack and its surroundings at variable angles between the plane of the crack and the direction of the X-rays.

example of this for a crack with $h/b = 10$ in aluminium. For heavier metals the decrease in contrast is even greater. The angle α must not, therefore, by greater than 5 to 10° if the crack is to be clearly visible in the photograph. Suspicious vague lines must for this reason always be photographed again at somewhat different angles. The close dependence of the contrast on the direction of the rays offers an excellent means of distinguishing cracks from corrosion grooves which often occur with old boilers. These much less dangerous grooves also give vague lines on the photograph.

ABSTRACTS OF RECENT SCIENTIFIC PUBLICATIONS OF THE N.V. PHILIPS' GLOEILAMPENFABRIEKEN

No. 1210: J. A. M. van Liempt and W. van Wijk: Die Löslichkeit von Krypton in verschiedenen Flüssigkeiten (Rec. Trav. chim. Pays Bas **56,** 632 - 634, June 1937).

The solubility of krypton in different liquids is determined. It is found to be much greater than that of argon. In pure glycerine krypton is only very slightly soluble: this liquid is therefore useful as a sealing liquid for krypton.

No. 1211: H. C. Hamaker: A system of colloid phenomena (Rec. Trav. chim. Pays Bas **56,** 727 - 747, June 1937).

Building further on the theoretical foundations given in **No. 1154** and **No. 1174** for the treatment of the lyophobe colloids, it is shown in this article that, as regards the attractive and repulsive forces between the particles, the possible properties of colloids may fall into two groups. These two groups of properties correspond to a large extent to the phenomena which are observed with lyohobe and lyophile colloids respectively. Theoretically however these two groups are not sharply divided, and the theory furnishes a clear description of the nature of the intermediate cases. Various results are in good agreement with experiment. Since the terms lyophobe and lyophile are in reality not applicable to the theoretical grouping, the writer proposes the distinguishing terms "reversible" and "irreversible". The classification into lyophobe and lyophile colloids can be retained at the same time. The two groupings are closely related but not identical. In this way a satisfactory classification of colloids and colloidal phenomena may be given.

No. 1212: J. H. Gisolf: Electron counters (Ned. T. Natuurk. **4,** 129 - 149, June 1937).

In this lecture, given before the Netherlands Physical Society, a survey was given of the action and the construction of discharge tubes for counting electrons.

No. 1213: Balth. v. d. Pol: A new theorem on electrical networks (Physica **4,** 585 - 589, July 1937).

With the aid of the modern form of Heaviside's symbolic calculus the following theorem is proved. If a constant unit EMF is applied to any given currentless network, the difference between the electrical and magnetic energy in the stationary final state is equal to half the derivative of the admittance to $j\omega$ for the limit $\omega = 0$. A special case of this is the well-known theorem that the efficiency with which a battery can charge a set of condensers is always 50 per cent.

No. 1214: J. Alfter and W. J. Oosterkamp: Ein Vorschlag für den Begriff Filmgüte bzw. Gütefaktor des Aufnahmematerials (Fortschr. Röntgenstr. **55,** 609 - 612, June 1937).

In order to obtain the same contrasts in the image of an object on different film material, in X-ray photographs of the lungs, the voltage of the tube must be chosen proportional to the gradation of the film at a blackening of 1. Only in relation to the gradation of the film does the tube voltage determine the character of the negative. As a standard of quality for X-ray film material a quantity is proposed which is proportional to the sensitivity and to the third power of the gradation of the film. This proposal refers to X-ray photographs of the lungs and is illustrated by photographs taken in practice.

No. 1215: Th. J. Weyers: Selectivity measurements of radio receiving sets (T. Ned. Rad.-Genoot. **7,** 156 - 172 July 1937).

Measurements relating to the selectivity of radio receiving sets are carried out according to various methods. It is found that reliable selectivity curves are obtained only when in measuring the conditions of normal use are imitated as closely as possible by applying to the apparatus to be tested two signals, one a "desired" signal, to which the apparatus is tuned, and a "disturbing" signal which is modulated with music. With a given intensity of the "desired" signal it may be established by ear what is the permissible intensity of the "disturbing" signal, if no disturbance is to be heard during the pauses in the modulation of the desired signal. In this method of measurement the following features are taken fully into account: cross detection, cross modulation, combination tone of desired and disturbing carrier wave and the suppression of the modulation of a weak signal by a simultaneous strong signal.